

卷之三

,Aro



132081

खण्ड 18 अंक 1, 2004
Vol. 18 No. 1, 2004

प्राकृतिक एवं भौतिकीय विज्ञान
शोध पत्रिका

JOURNAL OF NATURAL
&
PHYSICAL SCIENCES



गुरुकुल कांगड़ी विश्वविद्यालय, हरिद्वार
Gurukula Kangri Vishwavidyalaya, Hardwar

प्राकृतिक एवं भौतिकीय विज्ञान शोध पत्रिका

Journal of Natural & Physical Sciences

शोध पत्रिका पटल

JOURNAL COUNCIL

अध्यक्ष	स्वतन्त्र कुमार कुलपति	President	Swatantra Kumar Vice Chancellor
उपाध्यक्ष	एस. एल. सिंह प्राचार्य	Vice President	S. L. Singh Principal
सचिव	ए०के० चोपड़ा कुलसचिव	Secretary	A.K. Chopra Registrar
सदस्य	जय सिंह गुप्ता विज्ञाधिकारी एस. एल. सिंह* प्रधान सम्पादक जे. पी. विद्यालंकार व्यवसाय प्रबन्धक पी.पी. पाठक प्रबंध संपादक	Member	J. S. Gupta Finance Officer S. L. Singh* Chief Editor J. P. Vidyalankar Business Manager P. P. Pathak Managing Editor

सम्पादक मण्डल

वीरेन्द्र अरोड़ा (गणित)
बी० डी० जोशी (जन्तु विज्ञान)
आर. डी. कौशिक (रसायन शास्त्र)
वी. कुमार (कम्प्यूटर विज्ञान)
डी. के. माहेश्वरी (वनस्पति विज्ञान)
पी. पी. पाठक (भौतिकी) प्रबन्ध सम्पादक
एल.पी. पुरोहित (भौतिकी) सहायक सम्पादक
एस० एल० सिंह* (गणित) प्रधान सम्पादक

वाह्य सम्पादक

योल जे चो (ज्योगसांग नेशनल विविद, कोरिया)
एस. एन. मिश्रा (यूनिटरा, ड्र. अफ्रीका)
सोम नैम्पल्ली (इमरिट्स प्रोफेसर, कनाडा)
एस. पी. सिंह (अमेरियन विवि., कनाडा)
देवकी एन. तलवार (इण्डियाना विवि., यू.एस.ए.)

EDITORIAL BOARD

Virendra Arora (Mathematics)
B. D. Joshi (Zoology)
R. D. Kaushik (Chemistry)
V. Kumar (Computer Science)
D. K. Maheshwari (Botany)
P. P. Pathak (Physics) Managing Editor
L.P. Purohit (Physics) Asstt. Editor
S. L. Singh* (Mathematics) Chief Editor

Foreign Editors

Yeol Je Cho, E-mail : yjcho@nongae.gsnu.ac.kr
S. N. Mishra, E-mail : mishra@getafix.utr.ac.za
Som Naimpally, E-mail : sudha@accglobal.net
S. P. Singh, E-mail : spsingh@math.mun.ca
Devaki N. Talwar, E-mail : talwar@iup.edu

* Retiring on 31 January 2004, Prof. Virendra Arora will take over as Chief Editor.
CC-0. In Public Domain. Gurukul Kangri Collection, Hanwar

CONTENTS

CHEMISTRY

49-56 R.D. KAUSHIK, GARIMA SHARMA & CHARU ARORA : Antifungal activity of certain plants against *alternaria brassicae* and *aspergillus niger*.

115-124 R.D. KAUSHIK, SUREKHA KANNAUJIA AND SHASHI : Kinetics and mechanism of mn (II) catalysed periodate oxidation of o-toluidine.

MATHEMATICS

1-10 G.P. SAMANTA : Stochastic analysis of a demographic model of urbanization.

11-30 G. S. PANDEY : Calendar systems in ancient India.

31-40 N.G. SARKAR : Stochastic model of a population growth with periodically varying parameters.

41-48 K.P.R. RAO, N. SRINIVASA RAO AND B.V.S.N. HARIPRASAD : Three fixed point results for three maps.

57-67 V.K. SHARMA AND YOGITA BANA : Some geometrical constructions from baudhayana sulba sutra.

69-102 VINOD MISHRA AND S. L. SINGH : Similarity of plane figures and geometric & group theoretic study of cyclic quadrilaterals of first kind.

103-114 R.C. DIMRI AND V.B. CHANDOLA : A common fixed point theorem in fuzzy metric spaces.

132081

INSTRUCTIONS TO AUTHORS

This multidisciplinary journal is devoted to publishing research findings mainly in Biology, Chemistry, Physics & Mathematical Sciences. Two copies of good quality typed manuscripts (in Hindi or English) should be submitted to the Chief Editor or Managing Editor. Symbols are to be of the exactly same form in which they should appear in print. The manuscripts should conform the following general format : First page should contain (in the following order) title of the paper, Abstract (in English only), Keywords and phrases along with Subject Classifications, Name(s) of the author(s) with affiliation. Second page should start with main text with usual subheadings such as Introduction, Theory/material and methods, Results, Discussions etc. without numbering. Acknowledgement and References should be at the end. References should be quoted in the text as numbers in square brackets and grouped together at the end of the manuscript in the alphabetical order of the surnames of the authors. Abbreviations of journal citations should conform to the style used in the World List of Scientific Periodicals. Use double spacing throughout the manuscript. Here are some examples of citations in the references list :

S. A. Naimpally and B. D. Warrack : Proximity Space, Cambridge Univ. Press, U.K., 1970 (for books)

B. E. Rhoades : A comparison of various definitions of contractive mappings, Trans. Amer. Math. Soc. 226 (1977), 267-290 (For articles in journals, title of the articles are not essential in long review/survey articles.)

MANUSCRIPT SHOULD BE SENT TO : S. L. Singh, Chief Editor or P. P. Pathak, Managing Editor, JNPS, Science College, Gurukula Kangri Vishwavidyalaya, Hardwar -249404, India. *Email Address* (S.L.Singh): vedicmri@sancharnet.in

REPRINTS : Twenty five free reprints will be supplied. Additional reprints may be supplied at printer's cost.

EXCHANGE OF JOURNALS : Journals in exchange should be sent to the Chief Editor or Librarian, Gurukula Kangri Vishwavidyalaya, Hardwar 249404 INDIA

SUBSCRIPTION : Each volume of the journal is currently priced at Indian Rs.500 in SAARC countries and US \$ 100 elsewhere.

COPYRIGHT : Gurukula Kangri Viswavidyalaya, Hardwar. The advice and information in this journal are believed to be true and accurate but the persons associated with the production of the journal can not accept any legal responsibility for any errors or omissions that may be made.

Journal of Natural & Physical Sciences, Vol. 18 (2) (2004)

CONTENTS

CHEMISTRY

47-54 R.D. KAUSHIK, YOGESH C. NAINWAL, VIRENDRA SINGH AND A.K. CHAUBEY : Kinetics and mechanism of Mn (II) catalysed periodate oxidation of o-chloro aniline in acetone-water medium

MATHEMATICS

1-6 B.G. PACHPATTE : A note on generalizations of midpoint inequality.

7-16 K.P.R. RAO AND N. SRINIVASA RAO : Common fixed point theorems for four self-maps on metric spaces by altering distances

21-34 ANITA SHARMA AND V.K. SHARMA : Methods for solving cubic equations

35-46 YOGITA BANA AND V.K. SHARMA : Construction of some important vedis from Śulba sūtra

55-76 B.P. MISHRA & S.K. SINGH : Strong summability of functions based on (D, α, β) (C, I, m) summability methods

89-96 VIRENDRA ARORA AND NIDHI HANNA : Kātyāyna Śulba sūtra

97-104 S.L. SINGH AND AMAL M. HASHIM : New coincidences and fixed points of reciprocally continuous and compatible hybrid maps

105-116 S.L. SINGH AND RITU ARORA : Coincidence and fixed point theorems for hybrid contractions

PHYSICS

17-20 VIJAI KUMAR, R.P. VATS AND P.P. PATHAK : Harmful bioeffects of high frequency electromagnetic fields.

77-88 N. SINGH, DEVENDRA SINGH, VIKAS MISHRA AND PRAKASH MISHRA : Effect of polarizability on nucleation phenomenon during ice glaciation due to external electric field

Printed at : Kiran Offset Printing Press, Kankhal-Hardwar, Phone : +91-1334-245975

The manuscript should be typed in double spacing on both sides of A4 size white paper leaving wide margins. The manuscript should be arranged as follows : The first page should contain (in following order) Title, Abstract, Key-words & Phrases, Classification number, Authors' names and Address. The back of this page should remain blank. Next page should start with usual headings of Introduction, Material and methods/Theoretical formulation, Results, Discussion, Acknowledgement (if any) and References etc. In the text the the sentences like, " in previous paper (ref) the authors have....." or any such sentences which disclose the identity of authors, should not be used. References in the text are to be quoted by number in square brackets and are to be arranged alphabetically at the end. Thus the first reference appearing in the text would not necessarily be no.1. They are to be written as follows :

S. A. Naimpally and B. D. Warrack : Proximity Space, Cambridge Univ. Press, U.K., 1970
(for books)

B. E. Rhoades : A comparison of various definitions of contractive mappings, Trans. Amer. Math. Soc. 226 (1977), 267-290 (For articles in journals, title of the articles are not essential in long review/survey articles.)

Reference should not be made to unpublished work like Ph.D. thesis or papers communicated.

REPRINTS: Twenty five free reprints will be supplied. Additional reprints may be supplied at printer's cost.

EXCHANGE OF JOURNALS : Journals in exchange should be sent to the Chief Editor or Librarian, Gurukula Kangri Vishwavidyalaya, Haridwar 249404 INDIA

SUBSCRIPTION : Each volume of the journal is currently priced at Indian Rs.500 in SAARC countries and US \$ 100 elsewhere

COPYRIGHT : Gurukula Kangri Viswavidyalaya, Haridwar. The advice and information in this journal are believed to be true and accurate but the persons associated with the production of the journal can not accept any legal responsibility for any errors or omissions that may be made.

STOCHASTIC ANALYSIS OF A DEMOGRAPHIC MODEL OF URBANIZATION

G.P. SAMANTA

(Received 27.08.2002)

ABSTRACT

The continuous time version of Rogers' demographic model is considered. The demographic parameters describing the interactions are assumed to be represented by a dichotomic Markov process. Explicit expressions are derived for the time development of the average of the urban and rural population size and their asymptotic behaviours are discussed.

Mathematics subject classification (1980) : 92A17

Keywords : Demographic model, Dichotomic markov process, Urbanization, Stability, Delta-correlated Stochastic process.

INTRODUCTION

The study of long term growth and development of urbanization is essential for optimum provision of adequate and efficient public services in the urban sector and for the formulation and evaluation of public policy. Hence, the need for developing quantitative models of urbanization for improved population predictions. It is expected that the analysis of such models would facilitate the comprehension of the underlying mechanism of the phenomenon of urbanization. There have been several attempts to formulate demographic models for urban growth; important among them are the quantitative models by Keyfitz [1], Rogers [2] and the United Nations [3]. These three demographic models are discussed in detail in an article by Ledent [4].

The continuous time version of Rogers' demographic model (which was originally for discrete time) can be expressed by a pair of coupled linear differential equations [4]. Denoting by $U(t)$ the urban population and by $R(t)$ the rural population, this model becomes

$$\frac{dR}{dt} = (\alpha - m)R + \gamma U \quad (1a)$$

$$\frac{dU}{dt} = (\beta - \gamma)U + mR \quad (1b)$$

where α is the natural growth rate of the rural population, β is that of the urban population, m is the in-migration rate per individual of rural population (from rural to urban sector) and γ is the out-migration rate per urbanite (from urban to rural area).

In the present paper it is proposed to study the mean behaviour of the rural and urban population size for the demographic model described by equations (1) with demographic parameters stochasticized in order to take account of the effect of fluctuating environment. If we consider a physical situation in which the random environmental fluctuations do not occur with great rapidity in comparison with the time-scale of population growth, it is natural to assume that the stochastic parameters have finite correlation time.

MODIFIED STOCHASTIC DEMOGRAPHIC MODEL

The modified stochastic version of equations (1) to be considered is

$$\frac{d}{dt} X_1(t) = \beta_1(t)X_2(t) + \gamma_1(t)X_1(t) \quad (2a)$$

$$\frac{d}{dt} X_2(t) = \beta_2(t)X_1(t) + \gamma_2(t)X_2(t) \quad (2b)$$

where $X_1(t) = R(t)$, $X_2(t) = U(t)$, $\beta_1(t) = \gamma(t)$, $\beta_2(t) = m(t)$,

$\gamma_1(t) = \alpha(t) - m(t)$ and $\gamma_2(t) = \beta(t) - \gamma(t)$.

The stochastic parameters $\beta_i(t)$ and $\gamma_i(t)$ can be expressed as

$$\beta_i(t) = \beta_{io}[1 + \epsilon_i \Delta(t)], i = 1, 2 \quad (2c)$$

$$\gamma_i(t) = \gamma_{io}[1 + \epsilon_{i+2} \Delta(t)], i = 1, 2 \quad (2d)$$

where β_{io} and γ_{io} are constants and the ϵ_i 's are also constants but positive. $\Delta(t)$ is assumed to be a dichotomic Markov process (DMP) [5] defined as a two valued stepwise constant Markov process with equiprobable values +1 and -1 and transition probabilities (for $t_2 > t_1$):

$$P\{\Delta(t_2) = +1 | \Delta(t_1) = +1\} = P\{\Delta(t_1) = -1 | \Delta(t_1) = -1\}$$

$$= \frac{1}{2}[1 + e^{-v(t_2-t_1)}]$$

$$P\{\Delta(t_2) = +1 | \Delta(t_1) = -1\} = P\{\Delta(t_2) = -1 | \Delta(t_1) = +1\}$$

$$= \frac{1}{2}[1 - e^{-v(t_2-t_1)}]$$

Here $1/v$ is the correlation time.

For a DMP, we have

$$E[\Delta(t)] = 0, E[\Delta^2(t)] = 1,$$

$$E[\Delta(t_1)\Delta(t_2)] = e^{-v|t_1-t_2|} \quad (3)$$

Also the closure property holds exactly, i.e. if $\Psi\{\Delta(\cdot)\}$ is a functional of the process involving only times prior to t_1 , then

$$E[\Delta(t_1)\Delta(t_2)\Psi\{\Delta(\cdot)\}] = E[\Delta(t_1)\Delta(t_2)]E[\Psi\{\Delta(\cdot)\}] \quad (4)$$

AVERAGES OF THE POPULATIONS

Equations (2) can be written in matrix form:

$$\frac{d}{dt}X(t) = AX(t) + \Delta(t)BX(t)$$

$$\text{where } X(t) = \begin{bmatrix} X_1(t) \\ X_2(t) \end{bmatrix}, A = \begin{bmatrix} \gamma_{10} & \beta_{10} \\ \beta_{10} & \gamma_{20} \end{bmatrix}, B = \begin{bmatrix} \epsilon_3 \gamma_{10} & \epsilon_1 \beta_{10} \\ \epsilon_2 \beta_{20} & \epsilon_2 \gamma_{20} \end{bmatrix} \quad (5)$$

This has the solution

$$X(t) = e^{At}X(0) + \int_0^t e^{A(t-t')} \Delta(t') BX(t') dt'$$

Iterating,

$$X(t) = e^{At} X(0) + \int_0^t dt' e^{A(t-t')} \Delta(t') B [e^{At'} X(0) + \int_0^{t'} e^{A(t'-t'')} \Delta(t'') BX(t'') dt'']$$

Taking the ensemble averages, we get,

$$E[X(t)] = e^{At} X(0) + \int_0^t dt' e^{A(t-t')} B \left\{ \int_0^{t'} E[\Delta(t') \Delta(t'')] e^{A(t'-t'')} BX(t'') dt'' \right\} \quad (6)$$

Here we have used the result $E[(\Delta(t))] = 0$

$$\text{Therefore } E[X(t)] = e^{At} X(0) + \int_0^t e^{A(t-t')} B \mathcal{O}(t') dt' \quad (7)$$

$$\text{where } \mathcal{O}(t') = \int_0^{t'} e^{(A-vI)(t'-t'')} BE[X(t'')] dt'' \quad (8)$$

Thaking Laplace transform, (7) and (8) yield respectively

$$E[\bar{X}(s)] = [sI - A]^{-1} [X(0) + B\bar{\mathcal{O}}(s)]$$

$$\text{and } \bar{\mathcal{O}}(s) = [(s+v)I - A]^{-1} BE[\bar{X}(s)]$$

$$\text{whence } [sI - A - B[(s+v)I - A]^{-1} B] E[\bar{X}(s)] = X(0) \quad (9)$$

Now writing $D = (s+v-\gamma_{20})(s+v-\gamma_{10}) - \beta_{10}\beta_{20}$, it is seen that

$$[(s+v)I - A]^{-1} = \frac{1}{D} \begin{bmatrix} s+v-\gamma_{20} & \beta_{10} \\ \beta_{20} & s+v-\gamma_{10} \end{bmatrix}$$

Using the value of B from (5), we get

$$B[s+v]I - A]^{-1} B = \frac{1}{D} \begin{bmatrix} \mu_{11} & \mu_{12} \\ \mu_{21} & \mu_{22} \end{bmatrix},$$

where

$$\mu_{11} = \epsilon_3^2 \gamma_{10}^2 (s+v-\gamma_{20}) + \epsilon_1 \epsilon_3 \beta_{10} \beta_{20} \gamma_{10} + \epsilon_2 \epsilon_3 \beta_{10} \beta_{20} \gamma_{10} + \epsilon_1 \epsilon_2 \beta_{10} \beta_{20} (s+v-\gamma_{10}),$$

$$\mu_{12} = \epsilon_1 \epsilon_3 \beta_{10} \gamma_{10} (s + \nu - \gamma_{20}) + \epsilon_1^2 \beta_{10}^2 \beta_{20} + \epsilon_3 \epsilon_4 \beta_{10} \gamma_{10} \gamma_{20} + \epsilon_1 \epsilon_4 \beta_{10} \gamma_{20} (s + \nu - \gamma_{10}),$$

$$\mu_{21} = \epsilon_2 \epsilon_3 \beta_{20} \gamma_{10} (s + \nu - \gamma_{20}) + \epsilon_3 \epsilon_4 \beta_{20} \gamma_{10} \gamma_{20} + \epsilon_2^2 \beta_{10} \beta_{20}^2 + \epsilon_2 \epsilon_4 \beta_{20} \gamma_{20} (s + \nu - \gamma_{10}),$$

$$\mu_{22} = \epsilon_1 \epsilon_2 \beta_{10} \beta_{20} (s + \nu - \gamma_{20}) + \epsilon_1 \epsilon_4 \beta_{10} \beta_{20} \gamma_{20} + \epsilon_2 \epsilon_4 \beta_{10} \beta_{20} \gamma_{20} + \epsilon_4^2 \gamma_{20}^2 (s + \nu - \gamma_{10}),$$

$$\text{Writing } F = \left(s - \gamma_{10} - \frac{\mu_{11}}{D} \right) \left(s - \gamma_{20} - \frac{\mu_{22}}{D} \right) - \left(\beta_{10} + \frac{\mu_{12}}{D} \right) \left(\beta_{20} + \frac{\mu_{21}}{D} \right),$$

we find that

$$\{sI - A - B[(s + \nu)I - A]^{-1} B\}^{-1} = F^{-1} \begin{bmatrix} s - \gamma_{20} - \frac{\mu_{22}}{D} & \beta_{10} + \frac{\mu_{12}}{D} \\ \beta_{20} + \frac{\mu_{21}}{D} & s - \gamma_{10} - \frac{\mu_{11}}{D} \end{bmatrix} \quad (10)$$

From (9) and (10), we have

$$E[\bar{X}_1(s)] = F^{-1} \left[\left\{ s - \gamma_{20} - \frac{\mu_{22}}{D} \right\} X_1(0) + \left\{ \beta_{10} + \frac{\mu_{12}}{D} \right\} X_2(0) \right] \quad (11a)$$

$$\text{and } E[\bar{X}_2(s)] = F^{-1} \left[\left\{ s - \gamma_{10} - \frac{\mu_{11}}{D} \right\} X_2(0) + \left\{ \beta_{20} + \frac{\mu_{21}}{D} \right\} X_1(0) \right] \quad (11b)$$

$E[X_1(t)]$ and $E[X_2(t)]$ are respectively the inverse Laplace transform of (11), and their evaluation is laborious. Equations (11) give the values of $E[\bar{X}_1(s)]$ and $E[\bar{X}_2(s)]$ for different magnitudes of stochastic parametric perturbations.

DISCUSSION OF THE STABILITY OF A DELTA-CORRELATED STOCHASTIC PROCESS

Here we consider a physical situation in which the environmental changes occur with great rapidity as compared to the time-scale of population growth and consequently we take the random fluctuations due to them to be delta-correlated and the corresponding results can be obtained in the limit of infinitesimally short correlation time. To this end we let $\epsilon_i, \nu \rightarrow \infty$ such that $(\epsilon_i^2 / \nu) \rightarrow c_i$, $i = 1, 2, 3, 4$. Then the autocorrelation function

$$E[\in_i \Delta(t_1) \in_i \Delta(t_2)] \rightarrow 2c_i \delta(t_1 - t_2).$$

$$\text{Also } \frac{\mu_{11}}{D} \rightarrow c_3 \gamma_{10}^2 + \beta_{10} \beta_{20} \sqrt{c_1 c_2},$$

$$\frac{\mu_{12}}{D} \rightarrow \beta_{10} \sqrt{c_1} [\gamma_{10} \sqrt{c_3} + \gamma_{20} \sqrt{c_4}],$$

$$\frac{\mu_{21}}{D} \rightarrow \beta_{20} \sqrt{c_2} [\gamma_{10} \sqrt{c_3} + \gamma_{20} \sqrt{c_4}],$$

$$\frac{\mu_{22}}{D} \rightarrow c_4 \gamma_{20}^2 + \beta_{10} \beta_{20} \sqrt{c_1 c_2}.$$

Thus for a delta-correlated process,

$$E[\bar{X}_1(s)] = G^{-1} [\{s - \gamma_{20} - c_4 \gamma_{20}^2 - \beta_{10} \beta_{20} \sqrt{c_1 c_2}\} X_1(0) + \beta_{10} \{1 + \gamma_{10} \sqrt{c_1 c_3} + \gamma_{20} \sqrt{c_1 c_4}\} X_2(0)] \quad (12a)$$

$$E[\bar{X}_{20}(s)] = G^{-1} [\beta_{20} \{1 + \gamma_{10} \sqrt{c_2 c_3} + \gamma_{20} \sqrt{c_2 c_4}\} X_1(0) + \{s - \gamma_{10} - c_3 \gamma_{10}^2 - \beta_{10} \beta_{20} \sqrt{c_1 c_2}\} X_2(0)] \quad (12b)$$

where

$$G = \{s - \gamma_{10} - c_3 \gamma_{10}^2 - \beta_{10} \beta_{20} \sqrt{c_1 c_2}\} \{s - \gamma_{20} - c_4 \gamma_{20}^2 - \beta_{10} \beta_{20} \sqrt{c_1 c_2}\} \\ - \beta_{10} \beta_{20} \{1 + \gamma_{10} \sqrt{c_1 c_3} + \gamma_{20} \sqrt{c_1 c_4}\} \{1 + \gamma_{10} \sqrt{c_2 c_3} + \gamma_{20} \sqrt{c_2 c_4}\}$$

(12) can be written as

$$E[\bar{X}_1(s)] = \frac{(s+a)X_1(0) + cX_2(0)}{(s+a)(s+b) + d} \quad (13a)$$

$$\text{and } E[\bar{X}_2(s)] = \frac{(s+a)X_2(0) + c'X_1(0)}{(s+a)(s+b) + d} \quad (13b)$$

where

$$a = -(\gamma_{10} + c_3 \gamma_{10}^2 + \beta_{10} \beta_{20} \sqrt{c_1 c_2}),$$

$$b = -(\gamma_{20} + c_4 \gamma_{20}^2 + \beta_{10} \beta_{20} \sqrt{c_1 c_2}),$$

$$c = \beta_{10} (\gamma_{10} \sqrt{c_1 c_3} + \gamma_{20} \sqrt{c_1 c_4} + 1),$$

$$c' = \beta_{20} (\gamma_{10} \sqrt{c_2 c_3} + \gamma_{20} \sqrt{c_2 c_4} + 1),$$

$$d = -cc'.$$

$$\text{Let } \alpha, \beta = -\frac{1}{2} [-(a+b) \pm \sqrt{(a-b)^2 - 4d}]$$

Taking the inverse Laplace transform of (13),

$$E[X_1(t)] = \frac{1}{\alpha - \beta} [\{(\alpha + b)X_1(0) + cX_2(0)\}e^{\alpha t} - \{(\beta + b)X_1(0) + cX_2(0)\}e^{\beta t}] \quad (14a)$$

$$E[X_2(t)] = \frac{1}{\alpha - \beta} [\{(\alpha + a)X_2(0) + c'X_1(0)\}e^{\alpha t} - \{(\beta + a)X_2(0) + c'X_1(0)\}e^{\beta t}] \quad (14b)$$

Let us assume that the stochastic perturbation terms are the same. Hence $c_1 = c_2 = c_3 = c_4$. Now, $K^2 = (a-b)^2 - 4d = [1 + c_1(\gamma_{20} + \gamma_{10})]^2 \{(\gamma_{20} - \gamma_{10})^2 + 4\beta_{10}\beta_{20}\} > 0$ (15)

For asymptotically large times, $E[X_1(t)]$ and $E[X_2(t)]$ are unstable for $-(a+b) + K > 0$, i.e. $\gamma_{10} + \gamma_{20} > -c_1(\gamma_{10}^2 + \gamma_{20}^2 + 2\beta_{10}\beta_{20}) - K$ (16)

If $\gamma_{10} + \gamma_{20} < -c_1(\gamma_{10}^2 + \gamma_{20}^2 + 2\beta_{10}\beta_{20}) - K$, then $E[X_1(t)]$, $E[X_2(t)]$, will tend to zero, i.e. they will extinct after a long time span.

For $\gamma_{10} + \gamma_{20} = -c_1(\gamma_{10}^2 + \gamma_{20}^2 + 2\beta_{10}\beta_{20}) - K$, (17)

$$\lim_{t \rightarrow \infty} E[X_1(t)] = \frac{1}{K} \{(\alpha + b)X_1(0) + cX_2(0)\} \quad (18a)$$

$$\lim_{t \rightarrow \infty} E[X_2(t)] = \frac{1}{K} \{(\alpha + a)X_2(0) + c'X_1(0)\} \quad (18b)$$

In this case, we have

$$\lim_{t \rightarrow \infty} \frac{E[X_1(t)]}{E[X_1(t) + E[X_2(t)]]} = \frac{(\alpha + b)X_1(0) + c'X_2(0)}{\Delta} \quad (19a)$$

and

$$\lim_{t \rightarrow \infty} \frac{E[X_2(t)]}{E[X_1(t) + E[X_2(t)]]} = \frac{(\alpha + a)X_2(0) + c'X_1(0)}{\Delta} \quad (19b)$$

$$\text{where } \Delta = (\alpha + a + c)X_2(0) + (\alpha + b + c')X_1(0)$$

After some simple calculations, we have

$$\alpha + a = \{1 + c_1(\gamma_{20} + \gamma_{10})\} \left[\left(\frac{1}{2}(\gamma_{20} - \gamma_{10}) + \left\{ \frac{1}{4}(\gamma_{20} - \gamma_{10})^2 + \beta_{10}\beta_{20} \right\}^{1/2} \right) \right] \quad (20)$$

$$\alpha + b = \{1 + c_1(\gamma_{20} + \gamma_{10})\} \left[\left(\frac{1}{2}(\gamma_{10} - \gamma_{20}) + \left\{ \frac{1}{4}(\gamma_{20} - \gamma_{10})^2 + \beta_{10}\beta_{20} \right\}^{1/2} \right) \right] \quad (21)$$

If $\gamma_{20} > \gamma_{10}, \beta_{20} \geq \beta_{10}$ or $\gamma_{20} \geq \gamma_{10}, \beta_{20} > \beta_{10}$, then $\alpha + a > c$ and $\alpha + b > c'$,

$$\text{therefore } \frac{(\alpha + b)X_1(0) + c'X_2(0)}{\Delta} < \frac{(\alpha + a)X_2(0) + c'X_1(0)}{\Delta}$$

Therefore, we conclude that if mean value of $(\beta - \gamma) >$ mean value of $(\alpha - m)$ and mean value of $m \geq$ mean value of γ or mean value of $(\beta - \gamma) \geq$ mean value of $(\alpha - m)$ and mean value of $m >$ value of γ , then we have a high level of urbanization after a long time span.

If mean value of $(\beta - \gamma) =$ mean value of $(\alpha - m)$ and mean value of $m =$ mean value of γ , then the level of urban population and the level of rural population will be equal after a long time span.

It may be noted that for $\beta_{10} = 0$ (no migration from urban areas) Roger's model reduces to Keyfitz's model. For this model if the condition (17) holds good and mean value of $\beta - \gamma \geq$ mean value of $(\alpha - m)$, then

$$\lim_{t \rightarrow \infty} \left[\frac{E[X_1(t)]}{E[X_1(t)] + E[X_2(t)]} \right] = 0 \quad (22a)$$

$$\lim_{t \rightarrow \infty} \left[\frac{E[X_2(t)]}{E[X_1(t)] + E[X_2(t)]} \right] = 1 \quad (22b)$$

Therefore, in this case the entire population will eventually be urbanized.

CONCLUDING REMARKS

We have carried out an asymptotic analysis of a continuous stochastic version of Rogers' demographic model. It is seen that under some conditions this model holds up high level of urbanization. From this continuous stochastic version of Rogers' demographic model we have also deduced a continuous stochastic version of Keyfitz's demographic model here we have seen that under some conditions the entire population will eventually be urbanized.

ACKNOWLEDGEMENT

The author is grateful to Prof. Mrinal Miri, Director, Indian Institute of Advanced Study, Shimla and Dr. K.J. S. Chatrath, Secretary, Indian Institute of Advanced Study, Shimla for their help and encouragement.

REFERENCES

1. N. Keyfitz : Do cities grow by natural increase or migration? *Geographical Analysis* 12 (1980), 142-156.
2. A. Rogers : *Matrix Analysis of Interregional populations growth and distributions*, University of California Press, California (1968).

3. United Nations Population Division : Patterns of Urban and Rural population growth, New York (1980)
4. J. Ledent : Comparative dynamics of three demographic models of urbanization, IIASA Report, RR-80-1 (1980).
5. Prajneshu : A Stochastic model for two interacting species, Stochastic Processes and their Applications 4 (1976), 271-282.

CALENDAR SYSTEMS IN ANCIENT INDIA

G. S. PANDEY*

(Received 09-03-2003)

ABSTRACT

Ever since the dawn of Vedic civilization in India the counting of time was considered as the highest need of the society. During the Rgvedic period a solar year was recognized as formed of twelve months and each month of 30 days. Vedāṅga Jyotiṣa (1400 B.C.) introduced the concept of Yuga system of five saṃvatsaras (years) and made a clear distinction between a tithi (lunar day), Sāvana day and solar day. Also, in order to account for a solar year, the system of intercalary months was introduced.

In this exposition we, present a short account of the main fifteen calendar systems introduced in India from time to time. In particular, we trace the history of Vikram Samvata (initiated from 57 B.C.) formed on the basis of Surya Siddhānta, which is the only calendar system being used, with some variations, almost everywhere in India, while the Government of India has declared an alien, Saka Samvata (started from 78 A.D.), as the national calendar.

Mathematics Subject Classifications : 01 A 32, 85-03, 01 A 10.

Key Words & Phrases : Rigveda, Vedāṅga Jyotiṣa, Yuga, Tithi, Sāvana day, Solar day, Intercalary month, Kaliyuga, Surya Siddhānta, Vikrama Samvata and Saka Samvata.

INTRODUCTION

A calendar system is the first need of a civilized society. This is the reason that Ācārya Varāhmihira emphatically writes [Var b, Chapter II,8]:

“अप्रदीपा यथा रात्रिर् इनादित्यम् यथा नभः।
तथा असंवत्सरोराजा भ्रमन्ति इन्द्रइवध्वनिः॥”

That is, a king without calendar system is like a night without a lamp, the sky without the Sun and a blind man moving on the sound.

Ancient Indian scholars were keen observers of the celestial bodies and capable to explain a number of phenomena in the sky without any mechanical or optical instruments. The rise and setting of the Sun and Moon were carefully

recorded for the formulation of a suitable calendar. Also, it was recognized that the light of the Sun was composed of seven rays as in the following hymn of Rgveda [Rig, I, 16, 14, 2]:

“सप्त युज्जन्ति रथमेक चक्रमेको अश्वो वहति सप्तनाम।”

That is, the one-wheel chariot of (the Sun) is drawn by one horse bearing seven names.

During the Rgvedic period it was estimated that one year consisted of twelve months composed of three hundred sixty days and three hundred sixty nights as in the following couplet [Rig ; I, 164, 11]:

“द्वादशारं नाहिंताज्जराय वर्षर्ति चक्रं परिद्यामृतस्य।
आपुत्राऽग्ने मिथुना सो उत्र सप्त शतानि विंशतिश्च तस्युः॥”

That is, the Sun's wheel with twelve spokes (months) revolves around (the Earth) and is never destroyed. Oh Fire! on this wheel are mounted seven hundred and twenty people (360 days and 360 nights).

Vedic scholars recognized that the seasons were caused by the Sun [Rig ; I 95, 3]:

“पूर्वा मनुप्रादिशं पार्थिवाना मृतून प्राशा सद्विदधावनुष्टुः।”

That is, controlling the seasons (the Sun) creates the East at various places.

The above hymn indicates that the seasons are caused by the Sun and the eastern direction at every point is ascertained by the position of the Sun in the morning, which also suggests that the Earth is round and the local time differed at different places. Taittiriya Samhitā, on the other hand, asserts that the phases of the moon and its light are provided by the Sun [Tat; III, 41, 1]:

“चन्द्रमा उमावस्या आदित्य मनु प्रविशति।
सूर्य रश्मिश्चन्द्रमा गन्धर्व॥”

According to the climatic conditions of India, the whole year was divided into six parts and each part was called a *ritu* (ऋतु) as given in the following verse of Taittiriya Samhitā (cf.[Tat;4, 11]):

“मधुश्च माधवश्च वासन्तिकावृत् । शुक्रश्च शुचिश्च ग्रैष्मावृत् ॥
 नभश्च नभस्यश्च वार्षिकावृत् । इषश्चोर्जश्च शारदा वृत् ॥
 सहश्च सहस्यश्च हेमन्तिकावृत् । तपश्च तपस्यश्च शैशिरावृत् ॥”

That is, Madhu and Mādhava form the spring season, Śukra and Śuci the summer, Nabhaśa and Nabhaśya are the months of rainy season, Iṣa and Urja are Sarada, Sahasa and Sahasya form the Hemant, and Tapasa and Tapasya are the Sisira season.

SAMVATSARA AND ITS DIVISIONS

Taittirīya Samhitā provides the composition of a samvatsara in detail as in the following verses [tat ; V, 6, 7] :

“षड् रात्रीर्दीक्षितः स्यात् षड्वा क्रतवः संवत्सरः ।
 द्वादश रात्रीर्दीक्षितः स्यात् द्वादश मासः संवत्सरः ॥
 त्रयोदश रात्रीर्दीक्षितः स्यात् त्रयोदश मासः संवत्सरः ।
 पञ्चदश रात्रीर्दीक्षितः स्यात्पञ्चदश वा अर्धमासस्य रात्रयोर्धमासशः संवत्सर आप्यते ॥
 चतुर्विंशति रात्रीर्दीक्षितः स्याच्चतुर्विंशतिरर्ध मासः संवत्सरः ।
 त्रिशंतरात्रीर्दीक्षितः स्यात् त्रिशदक्षराविराट् - मासं दीक्षितः स्याद्यो मासः संवत्सरः ॥”

That is, be instructed in six nights because Samvatsara is of six seasons. Be instructed in twelve nights as samvatsara is of twelve months. Be instructed in thirteen nights as samvatsara is of thirteen months. Be instructed in fifteen nights as half-month has fifteen nights and half-months form samvatsara. Be instructed in twenty four nights, because samvatsara has twenty four half-months. Be instructed in thirty nights, thirty letters form Virat. Be instructed in a month for months form a samvatsara.

The above verse affirms that

1 Samvatsara = 6 seasons.

12 months = 24 half months.

13 months (including one intercalary month).

24 half - months.

As pointed out by Dikshit [Dik, pp. 42 - 43], the Vedic scholars regarded a lunar month of less than 30 days and, therefore, to account for a solar year it was necessary to add one intercalary month usually after 30 lunar months.

Although Nakshatras (constellations) are mentioned in Rgveda at a number of places, total list of 27 constellations are given in Taittiriya Śāmhīta (cf.(Tat) ; IV, 4/10 and, see also [Tat], 69 - 81, for some other details).

THE CONCEPT OF YUGA

During the Vedic and Vedāṅga periods the concept of five years Yuga was recognized. The year was called Samvatsaras and one Yuga consisted of five Samvatsaras. The first two verses of Rgvedic Jyotiṣa clearly affirms [cf. (Dik), p. 98] that:

“पञ्चसंवत्सरमयं युगाध्यक्षं प्रजापतिम्।
दिनत्वयनमासाङ्गं प्रणम्य शिरसा शुचिः॥ 1 ॥
प्रणम्य शिरसा कालमभिवाद्य सरस्वतीम्।
कालज्ञानं प्रवक्ष्यामि लगधस्य महात्मनः॥ 2 ॥”

That is to purify ourselves we pray to Prajāpati, the creator of Yuga composed of five Samvatsaras including days, seasons, ayans and months.

Praying to time and Saraswati (the goddess of learning), we describe the knowledge of time of Logadha.

The name of five samvatsaras are given in Taittiriya Brāhmaṇa in the form [Tat, III, 10/4] :

“संवत्सरोसि परिवत्स रोसि।
इदावत्सरोसि दुवत्सरोसि॥
इद्वत्सरोसिवत्सरोसि।.....”

The above lines give names of five Samvatsaras as Samvatsara, Parivatsara, Idavatsara, Iduvatsara and Idvatsara.

Vedāṅga Jyotiṣa provides further extension to Vedic calendar system. This classical text on Jyotiṣa contains 45 verses on ancient Indian astronomy. As mentioned above, Vedāṅga Jyotiṣa was composed by Lagadha around the year 1400 B.C. Dikshit (Dik, p. 94) has classified Vedāṅga Jyotiṣa into the following three sections.

- (i) R̄gveda Jyotiṣa.
- (ii) Yajurveda Jyotiṣa.
- (iii) Atharvaveda Jyotiṣa.

Of course, a number of verses of the above Jyotiṣa texts are similar. Vedāṅga Jyotiṣa provides the divisions and subdivisions of a Yuga in full detail as in the following couplet of Yajurveda Jyotiṣa. (cf. [Dik]; p 119, Yajurjyotiṣa Verse 27) :

“त्रिशत्यहनां सप्त षष्ठिरब्दः षड्क्रत्तबो इयने।
मासा द्वादशा सूर्याः स्युरेतत्पन्च गुणं युगम्॥”

That is, a Samvatsara consists of 366 days, six seasons, two ayans and twelve solar months. Yuga is its five times. Yajurjyotiṣa [loc. cit; p. 120, Verse 29] further specifies :

“उदयावासवस्य स्युर्दिन राशिः स्वपन्चकः।
क्रृषेद्द्वि षष्ठिहीनं स्यात् विशत्या चैकया स्तृणम्॥”

That is, in a Yuga the number of sunrise is five times the number of days in a solar year, while moon-rise is 62 less than that.

The above couplet asserts that in 1 Yuga the sun rises $5 \times 366 = 1830$ times, while the number of moon rise is $1830 - 62 = 1768$ times. It was known that the duration of time from one moonrise to the next was greater than one solar day. In fact, in one lunar month there can be 29 or 30 sunrise, while moonrise 28 or 29. the definition of a tithi depends on the motion of the Moon. It has been defined

as the time from a moon rise to the next. Savan day (अहोरात्र), on the other hand, is the time from a sunrise to the following, Yajurjyotiṣa makes a clear distinction between a tithi, Savan day and solar day as in the verse [Yajurveda Jyotiṣa, Verse 31]:

“सावनेन्दुस्तृमासानां सप्तिः सैका द्विसप्तिका।
द्युत्रिंशत् सावनः सार्धः सूर्यः स्तृणां सपर्ययः॥”

(In a Yuga) there are 61 Savan months, 62 luner months and 67 nakshatra months (स्तृमास). A savan month is of 30 days and a solar month of $30 \frac{1}{2}$ savan days.

Hence it is clear that a Yuga contains 60 solar months, 61 Savan months and 62 luner months. Since a Yuga contains 1830 days, a Savan month is of $1830 \div 61 = 30$ days. Also, one solar month = $1830 \div 60 = 30 \frac{1}{2}$ days. The following verse of Yajurjyotiṣa (loc. cit., Verse 37) provides the relationship between a Savan day, a solar day and a tithi and the intrinsic cause for the addition of an intercalary month to a luner calender system:

“द्यूनं द्विषष्टि भागेन हेयं सूर्यात् सपार्वणम्।
यत्कृतावुपजायेते मध्ये चान्ते चाधि मासकौ॥”

That is, by subtracting 62nd part from a Savan day a tithi (lunar day) is obtained, while adding up the 60th part to a Savan day it becomes a solar day. Since a Savan days is smaller than a solar day, the intercalary month occurs in the middle of end of a Yuga.

According to the above verse, we see that

$$\begin{aligned} 1 \text{ Tithi} &= \frac{1830}{62 \times 30} \quad \text{Savan days} \\ &= 61/62 \quad = 1 - 1/62 \text{ Savan days} \end{aligned}$$

and

$$1 \text{ solar day} = 30 \frac{1}{2} \div 30 \text{ Savan days}$$

$$= 61 / 60 = 1 + 1 / 60 \text{ Savan days.}$$

Hence, in order to account for a solar year, the addition of an intercalary month in a Yuga becomes necessary.

R̥gveda Jyotiṣa (R̥gjotisa), on the other hand, asserts that an intercalary month occurs after 30 Savan months as in the verse (cf. [Dik], p. 114) :

“यदर्थं दिनं भागानां सदा पर्वणि पर्वणि।
ऋतुं शेषं तु तद्विद्यात् संरव्याय सहपर्वणाम्॥ 23॥”

That is, the sum of half parts of tithis from all Parvas (half lunar months) forms a Rituśeṣa (remainder of lunar - month).

As mentioned above, one Yuga is formed of 1830 Savan days, 120 half - solar months and 124 half-lunar months (Parvas).

Hence we see That

$$\text{Half lunar - month (Parva)} = 1830 / 124 \text{ Savan days}$$

$$= 14 \frac{94}{124} \quad \text{Savan days,}$$

$$\text{Half solar-month} = \frac{1830}{120} \quad \text{Savan days}$$

$$= \frac{61}{4} \quad \text{Savan days}$$

$$= 15 \frac{1}{4} \quad \text{Savan days}$$

$$= 14 \frac{94}{124} + \frac{61}{124} \quad \text{Savan days}$$

Thus we have

$$(\text{Half solar - month}) - (\text{Half lunar - month})$$

$$= 15 \frac{1}{4} - 14 \frac{94}{124}$$

$$= \frac{61}{124} \quad \text{Savan days.}$$

This means, approximately half Savan days is the remainder. Therefore the difference in 30 Savan months (or 60 half Savan Months) becomes

$$\frac{60 \times 61}{124} = 29 \frac{64}{124} \quad \text{Savan days}$$

$$= 1 \text{ lunar - month.}$$

Hence, according to *Rgjotisa*, after 30 Savan months 1 lunar - months has to be added to account for solar years.

As mentioned above, the Savan days form the most basic elements for the formulation of all calendar systems on this planet, while lunar calendars are needed to ascertain the festivals and suspicious days of a year. Solar calendars, on the other hand, account for the periodic revolution of the Earth around the Sun. All these facts were fully understood and recognized during the *Vedāṅga* period.

COUNTING OF TIME

According to *Maṇusmṛti* (cf. [Dik], 148-149), the duration of Kali Yuga is 4,32,000 years and *Dvāpar*, *Treta* and *Kṛityuga* (*Satyuga*) are two, three and four times of Kaliyuga respectively. These four Yugas form a *MahāYuga* (Great-Yuga) which covers a time-span of 43,20,000 years, i.e., 10 times of Kali Yuga. One thousand *Mahāyugas* form a *Kalpa*, which is one day of *Brahmā*. From the beginning of the present *Kalpa* 453 *Mahāyugas* have elapsed and out of 454th *Mahāyuga* the first three Yugas, that is *kṛit*, *Treta* and *Dvāpar* have passed and at present it is Kaliyuga. Majumdar (cf. [Sh], p. 152) has estimated that Kaliyuga started from the year 3102 B.C. Whitney [see Dik, pp. 158-199] too has calculated that Kaliyuga begun from Thursday, the 17th February 3102 B.C., at midnight. According to *Sūryasiddhānta* too, the present year is the 5103rd year of Kaliyuga, which began from Thursday, Phalguna Krisna 15, at midnight. It may be mentioned here that the Sumerian Great Flood, which destroyed a greater part of the present day Middle East and North-Western Indian plains, occurred in the year 3100 B.C., i.e., about two years after the beginning of Kaliyuga. As a matter

of fact, there is no mention of this great flood in R̄gveda. But as pointed out by Sharma [Sh, p. 150], Atharvaveda has indicated it in connection with the availability of a medicinal plant in the form:

यत्र नावं प्रमंशनं यत्र हिमवन्तः शिरः।

That is, (the medicinal plant) occurs at the place where the boat glided down on the peak of Himavant (Himalaya).

Since the beginning of Kaliyuga has been mentioned in a number of scriptures, it may be considered as the initial point of the recorded calendar system in India. A calendar in Sanskrit is called "Paichāng" which means five parts, as it provides information mainly about Tithi (lunar-date), Vasar (day), Nakshatra, Yoga and Karan (Planetary positions). In this section we present a short introduction of some fairly known calendar systems of ancient India.

- (i) **Kaliyuga Era** : By adding up 3044 years to Vikram Saṁvat the Kaliyuga era of a time is obtained. At present Kaliyuga year begins from Caitra Śukla 1 like Vikram Saṁvat
- (ii) **Saptarśikāla** : It is known that Seven sages (सप्तर्षि) pass one constellation in 100 years and complete one round of zodiac in 2700 years. Saptarṣi era was based on the movements of seven sages in the sky. At the time of Alberuni it was used in "Kashmir and Punjab. It used to begin from caitra Śukla 1 like Vikram Saṁvat. At present it is not used anywhere in India.
- (iii) **Buddha Nirvāṇa Era** : It begins from Buddha Purnima; i.e., Vaisakha Sukla 15. This year it is 2546 Buddha Nirvana era. If we add 487 years to Vikram Saṁvat, we obtain Buddha Saṁvat.
- (iv) **Mahāvīra Nirvāṇa Era** : It begins from Kartik Sukla-1. By adding 470 years to Vikram Saṁvat, we get Mahāvīra Nirvāṇa Era.
- (v) **Vikram Saṁvat** : It is used in almost every part of India. It begins from caitra Śukla 1, but in Gujarat it begins from Kārtik Śukla 1 (Kartikadi). It

was started by the valiant king, Vikramāditya, of Malwa after defeating the Śaka hordes. By subtracting 57 years and 135 years from Vikram Saṃvat, we obtain Christian and Śaka eras respectively. Due to the great importance of Vikram Saṃvat, we shall discuss its history in detail in a separate section.

- (vi) **Christian Era** : Although initiated about 2000 year ago, this era was introduced by British rulers in India as the Official Calendar. It begins from 1st January and its year is solar. The month of February is of 28 days. January, March, May, July, August, October and December months are of 31 days, while April, June, September and November are of 30 days. It was started in Europe 57 years later than Vikram Saṃvat. Now a days it is used throughout India.
- (vii) **Śaka Era** : It was initiated 135 years later than Vikram Saṃvat to celebrate the coronation of Kuṣāṇ Emperor Kaniska (cf. [San], p. 203). It is still used in some parts of South India. The Government of India has declared it as the National Calendar. In all other respects it is similar to Vikram Saṃvat.
- (viii) **Cedi or Kalcuri Saṃvat** : It was started by the Kalcuri kings of Central India from Asvin Śukla 1 of Vikram Saṃvat 305. It was used in Central and Western part of India during the Kalcuri rule.
- (ix) **Gupta Era** : It was initiated to celebrate the coronation of the first Gupta emperor Chandragupta I, from Caitra Sukla 1 of Vikram Saṃvat 376 (319 A.D.). Chandragupta II was the most illustrious emperor of the Gupta dynasty. He defeated Śakas at the battle field of Karur near Multan (cf./ [Sec.], Vol. II, p. 6) and advancing to Vahlika region chased the mighty Huna hordes beyond Vāmkṣu river, which formed the northern frontier of India. Kalidas has given a vivid description of this battle fought on the south bank of Vāmkshu river in the allegorical form (Cf. [Kal.], Chapter IV, Verses 67 and 68):

विनीताध्वं श्रमास्तस्य वंक्षु तीरं विचेष्टनैः
 दुधुवर्वा जिनः स्कन्धान्लग्न कुकम केसरान्॥
 तत्र हूणावरोधानी भर्तृषु व्यक्त विक्रमम्।
 कपोल पाटलादेशि वभूवः रघु चेष्टितम्॥



That is, in order to shake off the fatigue (of the long march) the horses of Raghu rolled over the field and saffron filaments were seen on their shoulders. There the redness of the cheeks (by beating) of Huna queens, describing the bravery of their deceased husbands, testified the valour of Raghu.

It is said that after firmly securing the North-West frontiers of India against foreign invasions the Emperor, Chandragupta II, performed *Yajña* (यज्ञ) at the same place in Indraprastha where King Yudhishthira did *Rājsuya Yajña* for the recognition of his sovereignty.

After this *Yajña*, Chandragupta II assumed the title of "Vikramāditya" and actively promoted Vikram Samvat instead of Gupta era. Vikram Samvat, therefore, entirely superceded Gupta era. For other details and references see [Dik], P. 492.

Later on, during the reign of Emperor Budhgupta an Iron Pillar was established at the place of aforesaid *Yajna* under the supervision of Acharya Varahmihira, the greatest mathematician of that age, to commenorate the great achievements of Chandragupta II " Vikramaditya ". The inscriptions on the pillar is clearly readable of which one line is :

“तीत्वा सप्त मुखानि येन समरे सिन्धो र्जिता वाह्लीका॥”

That is, after crossing the river sindh with its seven tributaries, he (Chandragupta) conquered Vahlika (Bactria) region.

(x) **Harṣa Era** : It was initiated by Harṣavarddhana, a renowned king of Kannauj. A number of copper plates marked with this era has been obtained. By adding 663 years to Harṣa era we obtain Vikram Samvat. At the time of

Al Beruni it was being used around Kannauj and Mathura (cf. [Dik], p. 495 for other details).

- (xi) **Bengali Samvat** : It begins from Vaisakh Krisna 1 and by adding 650 years to it we get Vikram Samvat. At present Bengali era is 1409. It is used in Bengal.
- (xii) **Kollam Era** : It is used in Kerala. Begins from Sravana Śukla 1 and its year is solar. By adding 882 years to Kollam era, we obtain Vikram Samvat.
- (xiii) **Calukya Era** : It was started by Calukya King Vikramaditya from Vikram Samvat 1133. It was used in the Calukya kingdom upto Vikram Samvat 1219.
- (xiv) **Lakshman Sen Era** : It was started by the Sen King, Lakshman Sen, of Eastern Bihar and West Bengal from Vikram Samvat 1165. It was used in Central and Eastern Bihar from 13th to 17th Century of Vikram Samvat For other details see [Dik], pp. 496-497.
- (xv) **Jewish Era** : In order to save from Christian persention jewish people migrated to India during the early Christian era. Jewish settlements are found in some parts of Kerala, Maharastra and Gujarat. They use Jewish Calendar to ascertain their festivals and auspicious days. Jewish year begins towards the end of September and this year it is 5763rd of the Jewish era.

HISTORICITY OF VIKRAM SAMVAT

In order to count time, a number of calender systems were promulgated in India from time to time, but very few of them could survive beyond a few centruies. Older calendars like Kaliyuga Era (or Yudhisthira Samvat), and Maurya Samvat were used before the Vikram Samvat. Buddha Nirvan and Mahayira Nirvan Samvats are used in Buddhist and Jain Scriptures respectively. In the chronological order Vikram Samvat comes after Maurya Samvat. As mentioned in the preceding section, Vikram Samvat was initiated in the year 57 B.C. to

commemorate the great victory of King Vikramaditya against the śaka hordes.

Although some historians have doubted the existence of Vikramaditya and Vikram Samvat from 57 B.C., recently a number of research papers have been published to establish the historicity of the Malav King, Vikramaditya, and the Samvat initiated by him (cf. [Vyas a,b,c], [UPa], [UPb]. We give below the testimony of some impartial Muslim authors about the history of Vikramaditya and Vikram Samvat:

(i) **Alberni's Evidence :** It is well known in Indian history that about 450 years after king Vikramaditya of Malwa another valiant Emperor, Chandragupta II, of Gupta dynasty also assumed the title of Vikramaditya and defeated Saka hordes at the battle field of Karur near Multan. Alberuni has clearly distinguished the identity of these two heroes of Indian history (cf. [Sac], Vol. II, P. 6):

" The epoch of the era of śaka or śakala falls 135 years later than that of Vikramaditya. The here-mentioned śaka tyrannised over their country between the river Sindh and the ocean after he had made Aryavarta in the midst of this realm his dwelling place. The Hindus had much to suffer from him, till at last they received heep from the east, when Vikramaditya marched against him, put him to flight and killed him in the region of Karur, between Mustan and the castle of Loni. Now this date became famous, since there is a long interval between the era which is called the era of Vikramaditya and the killing of śaka, we think that the Vikramaditya from whom the era has got its name is not identical with that one who killed śaka, but only a name sake of his."

Abu Rihan Mahammed bin Ahmad al Biruni al Khwarizmi (970-1039 A.D.), who is known in India as 'Alleruni, was a great scholar and traveller. Abul Fazl Baihagi, who lived about half a century after Alberuni, writes in his famous treatise Tawarikh al-Subuktagini (cf. [ED], p.2):

" The Master Abu Rihan al Biruni excelled all his contemporaries in the science of philosophy, mathematics and geometry. Sultan of Khwarizm appointed

him to accompany the embassy which he sent to Mahmud of Ghazni."

Alberuni, following the army of Mahmud Ghazni, came to India and lived here for about ten years. He wrote an extensive account of his study in India in his famous book, *Kitab-u-Hindu*, which provides a vivid description of the contemporary Indian civilization, culture, science and technology. His testimony that Vikram Samvat was initiated by king Vikramāditya 135 years before the saka era provides a clinching argument to settle the problem.

(ii) **Evidence in Tabakat-I-Nasiri :** The author of *Tabakat-I-Nasiri*, Abu Umar Minhaj al Jurjani, was a native of Iran. He migrated to India during the reign of Slave dynasty and was appointed as Kazi of Delhi by Sultan Nasiruddin. In *Tabakat-I-Nasiri*, Jurjani (Juzjani) has given an account of the expedition of Sultan Altamash to conquer Malwa province. He writes (cf. [ED], vol. II, p. 328) :

In 632 H (1234 A.D.) Attamash lead an army against Malwa and took the city and fort of Bhelsa (Vidisha). There was a temple there which was three hundred years in building. It was 105 gaj high. It was demolished from Bhelsa he (Altamash) proceeded to Ujjain where there was the temple of Mahakala which he destroyed as well as an image of Vikramaditya who ruled Malwa 1346 years before this time. The Hindu era dates from his reign (as this figure leads to 1289 A.D. as the date of the capture of Ujjain which really is 1234 A.D. we may take 55 years as the length of Vikrama's rule). Some other images cast in copper were carried to Delhi with the stone image of Mahakala".

The above description of Jurjani implies that Vikramaditya ascended the throne of Malwa in the year (1234-1346) B.C. = 112 B.C. and initiated Vikram Samvat in the year 57 B.C. It may be mentioned here that the image of Vikramaditya carried away by Altamash from Mahakala temple was not of Chandragupta II 'Vikramaditya', because the latter was a Vaisnava.

(iii) Evidence in Bābur-Nāmā. According to William Erskine (Cf. [Bab], p. 715):

"Zahīrud-din-Muhammad Bābur was undoubtedly one of the most illustrious men of his age, and one of the most eminent and accomplished princes that ever adorned an Asiatic throne."

Bābur (1483-1530 A.D.) wrote his autobiography in Turkish language. He has given a vivid and extensive description of India including its history, geography, flora and fauna. About the history of Vikramaditya and Vikram Samvat Bābur, in 1527, categorically writes (cf. [Bab], p. 79):

"Not more than seven or eight observatories seem to have been constructed in the world. Māmūm Khalīfā made one with which the Māmūnī Tables were written. Batalumūs (Ptolemy) constructed another. Another was made in Hindustan, in the time of Rājā Vikramaditya Hindu, in Ujjain and Dhar, that is, the Malwa country, now known as Mandu. The Hindus of Hindustan use the Tables of this Observatory. They were put together 1584 years ago."

The above quotation from Bābur-Nāmā entails that two Observatories - one at Ujjain and another at Dhar-were constructed in the time of 'Raja Vikramaditya in the year 57 B. C. and a calendar (पंचाङ्ग) was formed on the basis of the data collected from the Observatories of Ujjain and Dhar established by him.

Mahapandit Rahul Sankrityayan, tracing the early history of Śakas in India (cf. [San], pp. 154-155), writes :

"मोग" (Mauves) की मृत्यु 58 ई. पू. के बाद भारत में शक राज्य छिन्न-भिन्न हो गया। (यह) विक्रम संवत का आरम्भ समय था।"

It is well known that Mog (or Maues) was a powerful Śakas ruler (77-58 B.C.), who conquered a substantial part of North India including Saurashtra, Ujjayini and Mathura. As pointed out by Rahul Sankrityayan, Mog died in the year 58 B.C. and thereafter the people of Malwa under the command of their valiant leader, Vikramaditya, defeated the Śakas hordes and liberated North India from Śakas tyranny. For other details see [UPa], pp. 16-62.

THE BASIS OF VIKRAM SAMVAT

Varahmihira, in Pañcasiddhāntikā, has given a systematic exposition of all five siddhantas prevalent at his time, namely Paitāmaha, Vasistha, Romaka, Pauliśa and. Sūrya Siddhānta Varahmihira assigns the first place to Sūrya Siddhānta in order of merit. He writes (cf. [Var a]; I,4)

पौलिशकृतः स्फुटोऽसौ तस्य सन्नस्तु रोमक प्रोकृः।
स्पष्टतरः सावित्रः परिशेषौ दूर विप्रष्टौ॥

That is, the Siddhānta composed by Pauliśa is accurate, Romaka is near to it, more accurate is Sāvitrah Sūrya Siddhānta), while the remaining two are far from the truth.

Due to the great importance of Sūrya Siddhānta in astronomical calculations, it has been preserved for more than 2000 years, while the subject matter of other four Siddhāntas are available only in their abridged form presented by Varāhmihīra in his Pañcasiddhāntikā. However, Sūrya Siddhānta known at the time of Varāhmihīra was somewhat different from the one available at present. Thibaut, for instance in his commentary on Pañcasiddhāntikā writes (cf. [Var a], p XIII):

“A cursory survey of those chapters of the Pañcasiddhāntikā which treat of the Sūrya Siddhānta shows at once that the treatise of that name known to Varāhmihīra agrees with the modern Sūrya Siddhānta in its fundamental features. The methods of the two treatises are essentially the same and on the other hand, sufficiently different from those of the other Siddhāntas summarized by Varāhmihīra, to ensure to the Sūrya Siddhānta in its two fold form a distinct position of its own. At the same time we can not fail to notice that in certain points the teaching of the old Sūrya Siddhānta (by which name I shall, for shortness sake, designate the Sūrya Siddhānta known to Varāhmihīra, must have differed from the corresponding doctrines of its modern representative.

Sūrya Siddhānta was composed completely on the basis of traditional Indian astronomy, being different in certain features from Paulisa and Romaka

Siddhāntas. Brahmagupta, in fact, has blamed Romaka Siddhānta for not following the traditional Hindu methods in some respects (for details see [Vara], p. LX VIII). Although some authors point out the influence of greek astronomy, Thibaut, on the other hand, emphasizes : [loc. at., pp. LXXIV-LXXV]:

“A work of the class of the Sūrya Siddhānta should spring directly from a work such as the Syntaxis, would be almost incomprehensible.....The merit of (its) originality, as far as it goes, would most probably belong to the unknown author of the old Sūrya Siddhānta.”

From the above quotation it is clear that Sūrya Siddhānta presents an outgrowth of traditional Indian Siddhāntas without any direct link with the Greek astronomy. Although Thibaut attributes the composition of Siddhāntas to an “unknown author”, in the beginning of modern Sūrya Siddhānta a reliable introduction of its original author is given in the form (cf. [SS], I, Coplets 2-4):

अल्पावशिष्टं तुकृते मयनामा महा सुरः।
रहस्यं परमं पुण्यं जिज्ञासुर्जानं मुत्तमम्॥२॥
वेदांगम् ग्रथम् इत्विलं ज्योतिषां गतिकारणम्।
आराध्ययन्वितं स्वन्तं तपस्ते पेसुदुश्चरम्॥३॥
तोषितस्तपसातेन प्रतिस्तस्मै वरार्थिमने।
ग्रहाणां चरितं प्रादान्मयाय सविता स्वयम्॥४॥

The above coplets state that when a small (part) of Kṛit(Saṃvat) remained a great Asur, named Maya, inquisitive of the meritorious (holy) secrets of Jyotiṣa (astronomy), which is the foremost part of Vedāṅga, studied the Sun with inaccessible ascetic fervour. The Sun (God) satisfied with his boon-aspiring asceticism provided him the characters of the planets.

According to Thibaut (cf. [Vara], p.LXXXIII), the Romaka Siddhānta was not composed later than 400 A.D. Dikshit, on the other hand, using some astronomical calculations (cf. [Dik], p 221) has inferred that the original Romaka Siddhānta was composed after Hipparchus (150 B.C.), but before the time of Ptolemy (150 A.D.). He further asserts (loc. cit., p.222) that Sūryasiddhānta was

composed earlier than Romaka. (cf.[Vyd], pp. 170-172).

He has mentioned that during an excavation of old mounds at Ujjain in the year 1938-1939 an old coin marked Maya (मय) was obtained.

Vyas writes that Mayasur was a resident of Ujjain, who composed Sūryasiddhānta to rectify a number of defects in the traditional Indian Siddhantas in the first century before the Christian era. Prior to 57 B.C. an old calendar, named Krit Samvat, was used in Malwa and a number of inscriptions and coins marked, Krit or / and Malav have been obtained. Vyas maintains that at the behest of King Vikramaditya, Mayasur thoroughly revised the old Krit Samvat and formulated a new one called 'Vikram Samvat'. Although Krit Samvat was in use for a few centuries more, Vikram Samvat took its place in due course. C.V. Vaidya (cf.[Vy d], p.39) too supports this assertion and claims that Vikram Samvat has been formulated on the basis of Sūryasiddhānta. It may be mentioned have that Varāhmihīra has included the name of Maya among the astronomers of ancient India (cf. [vy d], p. 172 in the form :

मय - यवत - मणित्थ शक्तिपूर्वे
दिवस करादिषु वासराः प्रदिष्टाः ॥

EPILOGUE

The science of astronomy was cultivated in India from the time of Rgveda. The Vedic scholars discovered that the Sun light was composed of seven rays, and one year consisted of 12 month and each month was of 30 days. A total list of 27 Nakshatras (constellations) are given in Taittiriya Samhitā. The concept of a Yuga consisting of five Samvatsaras (years) was recognized during the Vedāṅga period. The name of these Samvatsaras are given in Taittiriya Brahman Vedāṅga Jyotisa, which was composed by Lagadha around 1400 B.C., provides a more advanced aspect of the ancient Indian calendar. Yajur Jyotiṣa for instance, asserts that a Samvatsara consists of 12 solar months, six seasons and 366 days, and a Yuga is five times of a Samvatsara. Vedanga Jyotiṣa has clearly made distinction between the Savana solar and lunar months.

Manuśmṛti on the other hand, has given the concept of a Mahāyuga consisting of Krita, Treta, Dvāpar and Kaliyuga. According to Sūryasiddhānta the present year is the 510 3rd year of Kaliyuga which had began from Thursday, Phālguna Krisṇa 15, at midnight, that is, Kaliyuga had started from 17th February 3102 B.C. at midnight.

Although 15 calendar systems were introduced in ancient India from time to time, only Vikram Saṃvata and Christian era have survived. Vikram Saṃvata, which is based on the formulation of Sūryasiddhānta is being used with some variations in almost every part of India. It is really a great anomaly and totally improper that the Government of India has recognized alien Śakas Saṃvata as the National Calendar ignoring the claim of indigenous Vikrama Saṃvata.

REFERENCES

- [Bab] Bābur-Nāmā, Autobiography of Zahīr'd-din Muhammad Bābur (Translated by Annette S. Beveridge), Low Price Publications, B-2 Vardhaman Palace, Ashok Vihar IV, Delhi, 2000.
- [Dik] S.B. Dikshit, Indian Astronomy (Hindi Edition). Uttar Pradesh Hindi. Sansthan. Mahatma Gandhi Marg. Lucknow, 1990.
- [Div] H.N. Dwivedi, The Problem of Vikram in Indian History In : Saṃvat Pravartak-Samvat Vikramāditya. Edited by R.S. Vyas., Pandulipi Prakashan, 77/1 East Azad Nagar, Delhi, (1990), 99-127.
- [ED] S. Elliot and A Dowson, The History of India as Told by Its Historians, Low Price Publications, B-2, Vardhaman Palace, Ashok Vihar IV. Delhi, 1990.
- [Kal] Kalidas Raghuvamsha Mahakavyam, Chaukhamba Surbharati Prakashan, Varanasi, 1994.
- [L.S] Labshaman Swarup, Vikram ki Aitihasikata In: Saṃvat Pravartak-Samrat Vikramāditya, Edited by R.S. Vyas, Pandulipi Prakashan, 77/1 East Azad Nagar, Delhi, (1990), 84-98.
- [Pan] Raj Bali Pandey, Historicity of Vikram (Hindi) In: Saṃvat Pravartak-Samrat Vikramāditya, Edited by Raj Shekhar Vyas, Pandulipi Prakashan, 77/1 East Azad Nagar, Delhi, (1990), 128-141.
- [Reu] V.N. Reu, Vikram Saṃvata In : Saṃvat Pravartak-Samrat Vikramāditya

Edited by R.S. Vyas, Pandulipi Prakashan, 77/1 East Azad Nagar, Delhi (1990), 142-164.

- [Rig] Rigveda, Edited by Vasudeo Sharma and Krisna Bhatta, Gore. Nirnaya Sagar Press, Bombay, 1910.
- [Sac] E.Sachau, Alberuni's India, Vol. II, London, 1910.
- [San] Rahul Sankrityayan, History of Central Asia, Vol. I, Bihar Rastra Bhasa Parisad, Patna, 1985.
- [Sh] Ram Vilas Shara, Western Asia & Rigveda (Hindi), Hindi Madhyam Karyavaya Nideshalaya, University of Delhi, 1994.
- [SS] Sūrya Siddhānta (Edited by Baldev Prasad Mishra) Shri Venkateshwar Pess, Bombay, 1896.
- [Tat] Taittirīya Samhitā, Vedic Samsodhan Mandal, Poona, 1972.
- [Upa] B.S. Upadhyaya History of Vikram Samat In: Samvat Pravartak-Samrat Vikramaditya. (Edited by R.S. Vyas), Pandulipi Prakashan, 77/1 East Azad Nagar, Delhi, (1990), 52-83.
- [Upb] B.S. Upadhyaya, Feeders of Indian Culture, Peoples Publishing House, New Delhi, 1973.
- [Vara] Ācharya Varāhmihira, Pañcasiddhāntikā (Edited by G. Thibaut and Sudhakar Dvivedi) Kashi Sanskrit Series, Varanasi, 1889.
- [Var.b] Ācharya Varāhmihira, Brhat Samhitā, (Edited by Sudhakar Dvivedi), Vol. I & II, Kashi Sanskrit Series, Varanasi, 1895-1897.
- [Vy a] S.N. Vyas Vikramāditya-IIIusions about Existence and their Extripation, Anustup, M.P. Government Literary Council, Bhopal (1972), 1-34.
- [Vy b] S.N. Vyas, Samvat Pravartak-Samrat Vikramāditya (Edited by R.S. Vyas), Pandulipi Prakashan, 77/1, East Azad Nagar, Delhi (1990), 1-27.
- [Vy c] S. N. Vyas, Krit Samvat - Investigation of its Main Causes, Anustup, M.P. Government Literary Council, Bhopal (1572), 35-53.
- [Vy d] S.N. Vyas, Maya Culture and its Relation with Malva, Anustup, M.P. Government Literary Council, Bhopal (1572), 169-172.

STOCHASTIC MODEL OF A POPULATION GROWTH WITH PERIODICALLY VARYING PARAMETERS

N.G. SARKAR

(Received 30.05.2003 and in revised form 14.07.2003)

ABSTRACT

The paper deals with the study of the stochastic behaviour of a population obeying Gompertzian growth equation with both periodic growth coefficient and periodic carrying capacity under the influence of randomly fluctuating environment.

Mathematics subject classification (2000) : 92A17

Key words: Gompertz equation, Harmonic growth coefficient, White noise, Stochastic process, Periodic carrying capacity, Transfer function.

INTRODUCTION

The growth plays a significant role in life-process. There are different models of growth processes. The Malthusian [4] growth for which the growth coefficient α constant is meaningless after a large time t due to the ridiculously large value of the population size. An important generalisation of Malthusian model is the Gompertz model equation which finds wide application in different growth processes of men, animals, plants, tumour etc. [3]. Gompertz model equation is given by [1,2,3].

$$\frac{dN(t)}{dt} = -\alpha N(t) \ln \frac{N(t)}{K} \quad \dots \dots \quad (1.1)$$

with $N(0) = N_0$, $N(\infty) = K$.

Writing $x(t) = \ln N(t)$, the equation (1.1) reduces to

$$\frac{dx(t)}{dt} = -\alpha x(t) + \alpha \ln K \quad \dots \dots \quad (1.2)$$

where $N(t) \geq 1 \Rightarrow x(t) \geq 0 \quad \dots \dots \quad (1.3)$

In the above equation (1.1) or (1.2), α is the growth coefficient and K is the carrying capacity [1,2]. A great many phenomena are cyclic or oscillatory in behaviour. Due to diurnal or annual influences, there are numerous environmental factors like temperature, sunlight etc. which affect the growth processes. As a result cyclic variations occur in the growth process and the growth coefficient α and carrying capacity may vary periodically with time. Besides, the randomly fluctuating environment effects on the growth processes. The object of the present paper is to study the overall effects of the periodically varying growth coefficient or carrying capacity and the randomly fluctuating environment on the growth behaviours of a population obeying Gompertzian model equation.

GOMPERTZIAN GROWTH IN RANDOM ENVIRONMENT : SINUSOIDALLY VARYING GROWTH COEFFICIENT

We shall start with a simple harmonic growth coefficient and take [1].

$$\alpha(t) = a_0 + a \sin(\omega t + \varphi), a_0, a < 0 \quad \dots \dots \quad (2.1)$$

This says that the growth coefficient, $\alpha(t)$, has an average value, a_0 , and oscillates with amplitude, a , about this average value at a frequency, $\omega = 2\pi / T$; T is the period of oscillation and φ is the phase angle. The oscillation increases as a increases and no oscillation for $a = 0$. The phase angle φ is : $0, \pi/2, \pi, 3\pi/2$. For small oscillation in $\alpha(t)$ a and a_0 satisfy the condition:

$$|a/a_0| \ll 1. \quad \dots \dots \quad (2.2)$$

Putting equation (2.1) in equation (1.2) we get

$$\frac{dx(t)}{dt} + \alpha(t)x(t) = \alpha(t) \ln K = F_1(t) \quad (\text{say}) \quad \dots \dots \quad (2.3)$$

where $F_1(t)$ is given by

$$F_1(t) = \{a_0 + a \sin(\omega t + \varphi)\} \ln K \quad \dots \dots \quad (2.4)$$

To take account of the influence of the randomly fluctuating environment we add in equation (2.3) a white noise (i.e. fluctuating term) $f(t)$ whose properties are given by [8]

$$\langle f(t) \rangle = 0, \langle f(t)f(t') \rangle = 2D(t)\delta(t-t') \quad \dots \dots \quad (2.5)$$

where $\langle \rangle$ denotes the average over the ensemble of the stochastic process, $D(t)$ is the intensity (or diffusion coefficient) of the white noise. Then the equation (2.3) becomes

$$\frac{dx(t)}{dt} + \alpha(t)x(t) = F_1(t) + f(t) = \beta_1(t), \quad (\text{say}) \quad \dots \dots \quad (2.6)$$

where

$$\begin{aligned} \langle \beta_1(t) \rangle &= \langle F_1(t) + f(t) \rangle \\ &= F_1(t) \text{ and } \langle \beta_1(t)\beta_1(t') \rangle = F_1^2(t) + 2D(t)\delta(t-t') \end{aligned} \quad \dots \dots \quad (2.7)$$

The solution of the equation (2.6) can be written as [6]

$$x(t) = \int_0^t g_1(t, \tau) \beta_1(\tau) d\tau = \int_0^t g_1(t, \tau) [F_1(\tau) + f(\tau)] d\tau \quad \dots \dots \quad (2.8)$$

where the transfer function $g_1(t, \tau)$ is given by [6]

$$g_1(t, \tau) = 0, \tau < t, \quad g_1(t, \tau) = 1 \text{ and}$$

$$\frac{dg_1(t, \tau)}{dt} = -\alpha(t) \quad g_1(t, \tau) \quad \dots \dots \quad (2.9)$$

The equation (2.9) leads to

$$g_1(t, \tau) = e^{-\int \alpha(t) dt} = e^{-a_0 t + \frac{a}{\omega} \cos(\omega t + \varphi) + a\tau - \frac{a}{\omega} \cos(\omega\tau + \varphi)} \quad \dots \dots \quad (2.10)$$

The variance of the stochastic process $x(t)$ is

$$\begin{aligned} \sigma_1^2(t) &= \langle x^2(t) \rangle - \langle x(t) \rangle^2 \\ &= \iint_0^t g_1(t, \tau) g_1(t, \tau') \langle \beta_1(\tau) \beta_1(\tau') \rangle d\tau d\tau' \\ &= \int_0^t g_1^2(t, \tau) [F_1^2(\tau) + 2D(\tau)] d\tau \quad \dots \dots \quad (2.11) \end{aligned}$$

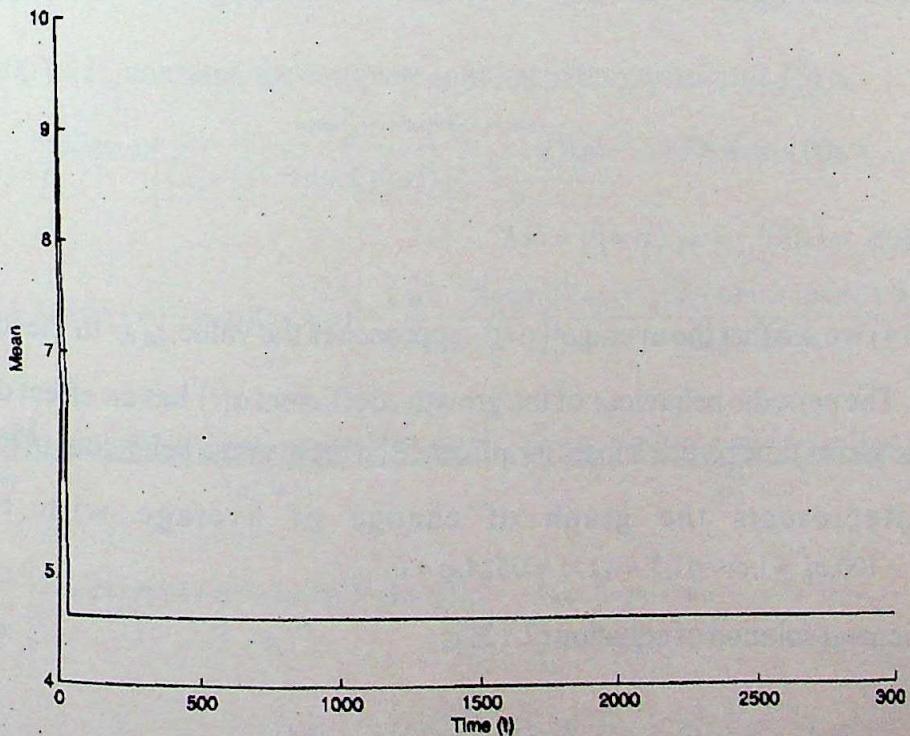


Fig. 1: Changes of averages $\langle x(t) \rangle_1$ with time t in days for sinusoidally varying growth coefficient.

Differentiating equation (2.11) w.r. to time t [7] we have

$$\frac{d\sigma_1^2(t)}{dt} + 2\alpha(t)\sigma_1^2(t) = F_1^2(t) + 2D(t) \quad \dots \dots \quad (2.12)$$

which is the basic equation of the time evolution of the variance of the stochastic process $x(t)$ and $F_1(t)$ is given by equation (2.4). The average value of $x(t)$ is $\langle x(t) \rangle$, then equation (2.6) gives

$$\frac{d\langle x(t) \rangle_1}{dt} + \alpha(t)\langle x(t) \rangle_1 = F_1(t) \quad \dots \dots \quad (2.13)$$

with initial condition $\langle x(0) \rangle_1 = x_0$ (say),

The complete solution of equation (2.13) is

$$\langle x(t) \rangle_1 = \langle x(t) \rangle_{1h} \int_0^t \left\{ \langle x(t) \rangle_{1h} \right\}^{-1} F_1(t) + \langle x(t) \rangle_{1h}$$

or,

$$\langle x(t) \rangle_1 = \ln K + (x_0 - \ln K) e^{-a_0 t + \frac{a}{\omega} \cos(\omega t + \varphi) - \frac{a}{\omega} \cos \varphi} \quad \dots \dots \quad (2.14)$$

where $\langle x(0) \rangle_1 = \langle x(0) \rangle_{1h} = x_0$, $\langle x(\infty) \rangle_1 = \ln K$

From (2.14) we see that the average $\langle x(t) \rangle_1$ approaches the value $\ln K$ in the long run as expected. The periodic behaviour of the growth coefficient $\alpha(t)$ has an effect only in the initial state and as time passes it loses its influence on the average behaviour of the system.

Fig. I Represents the graph of change of average with time for $x_0 = 10, K = 100, a_0 = 1, a = 0.1, T = 12, \omega = 0.524, \varphi = 0$.

General solution of equation (2.12) is

$$\sigma_1^2(t) = \sigma_{1h}^2(t) \int_0^t \left\{ \sigma_{1h}^2(t) \right\}^{-1} [F_1^2(t) + 2D(t)] dt + \sigma_{1h}^2(t) \quad \dots \dots \quad (2.15)$$

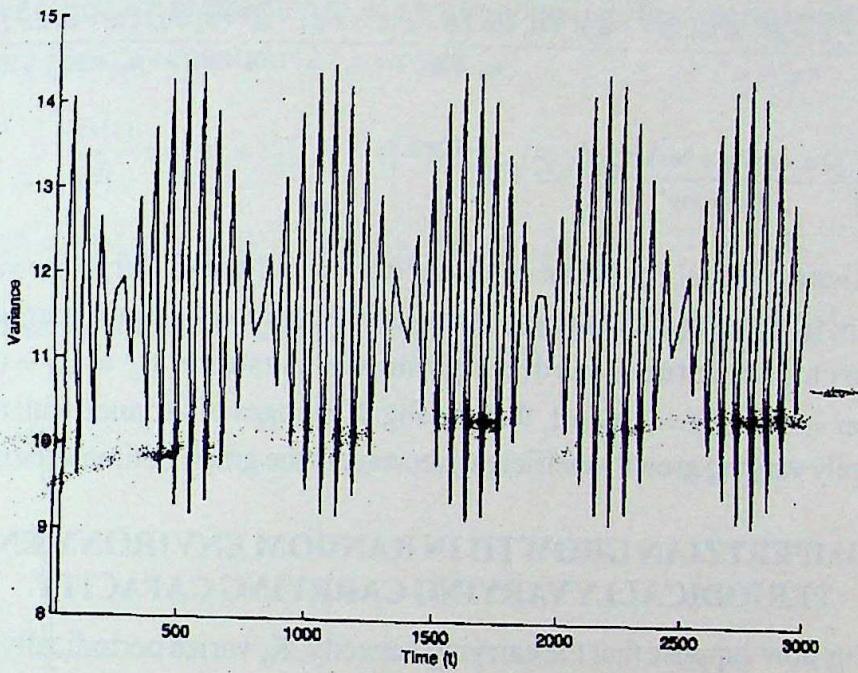


Fig. 2: Changes of variance $\sigma_2^2(t)$ with time t in days for sinusoidally varying growth coefficient.

Assuming $D(t) = D$, constant, the complete solution of the equation (2.15) is

$$\begin{aligned}
 \sigma_1^2(t) = & e^{\frac{2a\cos(\omega t + \varphi)}{\omega}} \left[\left\{ a_0^2 + (a^2/2)(\ln K)^2 2D \right\} \frac{1}{2a_0} + \frac{2a_0 a}{4a_0^2 + \omega^2} \times \right. \\
 & \times \left\{ 2a_0 \sin(\omega t + \varphi) - \omega \cos(\omega t + \varphi) \right\} - \left(\frac{a^3}{2\omega} + \frac{2a_0^2 a}{\omega} \right) \left\{ \frac{2a_0 \cos(\omega t + \varphi) + \omega \sin(\omega t + \varphi)}{4a_0^2 + \omega^2} \right\} - \\
 & - \frac{a_0 a^2}{\omega} \times \left\{ \frac{a_0 \sin 2(\omega t + \varphi) - \omega \cos 2(\omega t + \varphi)}{a_0^2 + \omega^2} \right\} - \frac{a^2}{4} \left\{ \frac{a_0 \cos 2(\omega t + \varphi) + \omega \sin 2(\omega t + \varphi)}{a_0^2 + \omega^2} \right\} + \\
 & + \frac{a^3}{2\omega} \left\{ \frac{2a_0 \cos 3(\omega t + \varphi) + 3\omega \sin 3(\omega t + \varphi)}{4a_0^2 + 9\omega^2} \right\} \left] - e^{-2a_0 t + \frac{2a}{\omega} \cos(\omega t + \varphi)} \times \right. \\
 & \times \left[\left\{ (a_0^2 + a^2/2)(\ln K)^2 + 2D \right\} \frac{1}{2a_0} + \frac{2a_0 a (2a_0 \sin \varphi - \omega \cos \varphi)}{4a_0^2 + \omega^2} - \left(\frac{a^3}{2\omega} + \frac{2a_0^2 a}{\omega} \right) \times \right.
 \end{aligned}$$

$$\begin{aligned}
 & \times \left(\frac{2a_0 \cos \varphi + \omega \sin \varphi}{4a_0^2 + \omega^2} \right) - \frac{a_0 a^2 (a_0 \sin 2\varphi - \omega \cos 2\varphi)}{a_0^2 + \omega^2} - \frac{a^2}{4} \frac{(a_0 \cos 2\varphi + \omega \sin 2\varphi)}{a_0^2 + \omega^2} + \\
 & + \frac{a^3}{2\omega} \frac{(2a_0 \cos 3\varphi + 3\omega \sin 3\varphi)}{4a_0^2 + 9\omega^2} - \sigma^2(0) e^{-\frac{2a_0 \cos \varphi}{\omega}} \quad \dots \dots \quad (2.16)
 \end{aligned}$$

which satisfies the initial condition $\sigma_1^2(0) = \sigma^2(0)$. When time t is large the variance (or fluctuation) $\sigma_1^2(t)$ remains finite and oscillatory with superpositions of a number of waves with different amplitudes and frequencies. Assume $a_0 = 1, a = 0.1, T = 12, \omega = 0.524, \varphi = 0, \sigma^2(0) = 2, D = 1$, then the Fig.2 (changes of variance with time in days for sinusoidally varying growth coefficient) represents the graph of time versus variance.

GOMPERTZIAN GROWTH IN RANDOM ENVIRONMENT: PERIODICALLY VARYING CARRYING CAPACITY

Let us now suppose that the carrying capacity, K , varies periodically with time t because of either diurnal or annual influences. The simplest possible example of carrying capacity oscillatory with angular frequency ω is obtained by assuming [5].

$$K(t) = K_0(1 + b \cos \omega t) \quad \dots \dots \quad (3.1)$$

where K_0 is the average value of K over a long time interval and provided

$$0 < b \ll 1 \quad \dots \dots \quad (3.2)$$

i.e., provided the fluctuations in K are small relative to K_0 . The equation (3.1) can be written as:

$$K(t) = K_0 e^{b \cos \omega t} \quad \dots \dots \quad (3.3)$$

provided the condition (3.2) is satisfied.

Due to periodic effect in carrying capacity $K(t)$ given by equation (3.1), the growing population size varies. Putting equation (3.1) in equation (1.2) taking growth coefficient α constant we get:

$$\frac{dx(t)}{dt} + \alpha x(t) = \alpha \ln K(t) = F_2(t), \quad (\text{say}) \quad \dots \dots \quad (3.4)$$

where

$$F_2(t) = \alpha(\ln K_0 + b \cos \omega t) \quad \dots \dots \quad (3.5)$$

As before we introduce in equation (3.4) a white noise (i.e., fluctuating term) $f(t)$ whose properties are given by equation (2.5), we get:

$$\frac{dx(t)}{dt} + \alpha x(t) = F_2(t) + f(t) = \beta_2(t), \quad (\text{say}) \quad \dots \quad \dots \quad (3.6)$$

$$\text{where } \langle \beta_2(t) \rangle = F_2(t), \langle \beta_2(t) \beta_2(t') \rangle = F_2^2(t) + 2D(t)\delta(t-t') \quad \dots \quad \dots \quad (3.7)$$

We can take suitable function $g_2(t, \tau)$, whose properties are same as $g_1(t, \tau)$, in section (2). Then

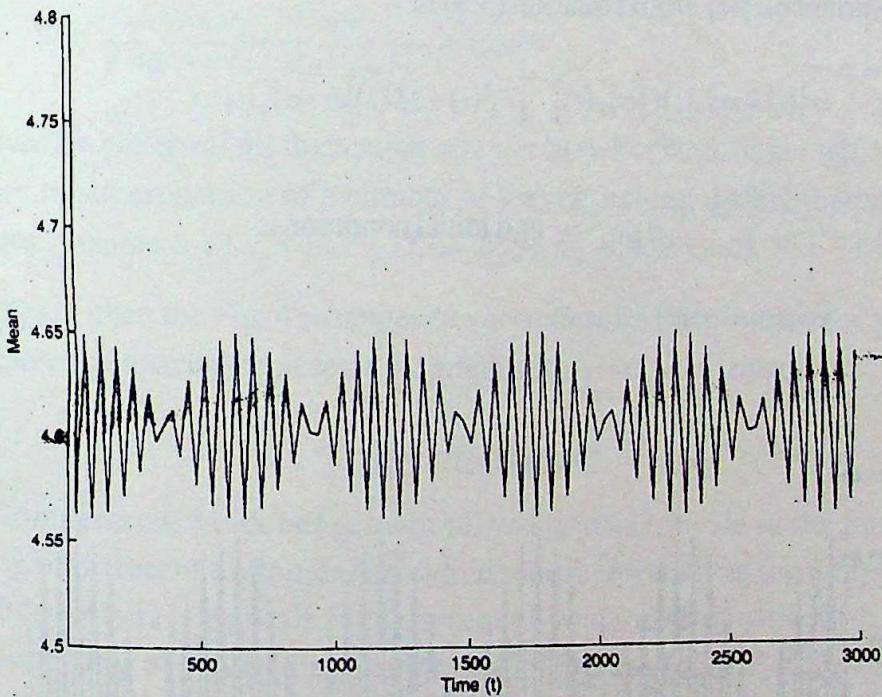


Fig. 3: Changes of average $\langle x(t) \rangle_2$ with time t in days for periodically varying carrying capacity.

$$g_2(t, \tau) = e^{-\int_{\tau}^t \alpha dt} = e^{-\alpha t + \alpha \tau} \quad \dots \quad \dots \quad (3.8)$$

According to the section (2) we get the variance equation for the stochastic process $x(t)$ where $F_2(t)$ is given by equation (3.5). The average value $\langle x(t) \rangle_2$ of $x(t)$ is given:

$$\frac{d \langle x(t) \rangle_2}{dt} + \alpha \langle x(t) \rangle_2 = F_2(t), \langle x(0) \rangle_2 = x_0, \quad (\text{say}) \quad \dots \quad \dots \quad (3.9)$$

The general solution equation (3.9) is

$$\langle x(t) \rangle_2 = \ln K_0 + \frac{\alpha b(\alpha \cos \omega t + \omega \sin \omega t)}{\alpha^2 + \omega^2} + \left(x_0 - \ln K_0 - \frac{\alpha^2 b}{\alpha^2 + \omega^2} \right) e^{-\alpha t} \dots \quad (3.10)$$

From (3.10) we see that the average $\langle x(t) \rangle_2$ remains finite always i.e., the population size $N(t)$ remains finite always and after a large time, the population shows periodic behaviour.

Fig.3. Represents the graph of time verses average $\langle x(t) \rangle_2$ for $x_0 = 10, \alpha = 1, \omega = 1, T = 12, \omega = 0.524, K_0 = 100, b = 0.05$. Now according to section (2) the solution for variance $\sigma_2^2(t)$ from equation (3.9) is

$$\sigma_2^2(t) = \sigma_{2h}^2(t) \int_0^t \left\{ \sigma_{2h}^2(t) \right\}^{-1} [F_2^2(t) + 2D(t)] dt + \sigma_{2h}^2(t) \dots \dots \quad (3.11)$$

where $\sigma_{2h}^2(t) = \sigma^2(0) e^{-\int_0^t 2\alpha dt} = \sigma^2(0) e^{-2\alpha t}$ and initial condition is

$$\sigma_{2h}^2(0) = \sigma^2(0), \text{ (say)} \dots \dots \quad (3.12)$$

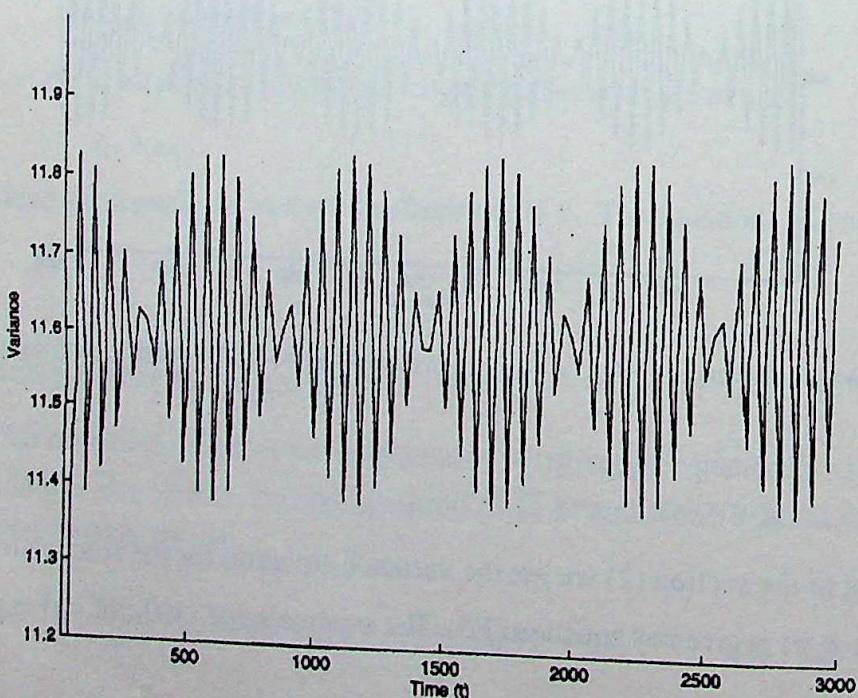


Fig. 4: Changes of variance $\sigma_2^2(t)$ with time t in days for periodically varying carrying capacity.

Assuming $D(t) = D$ constant, then equation (3.12) gives

$$\sigma_2^2(t) = e^{-2\alpha t} \left[\alpha^2 \int_0^t (\ln K_0 + b \cos \omega t)^2 e^{2\alpha t} dt + 2D \int_0^t e^{-2\alpha t} dt \right] + \sigma^2(0)e^{-2\alpha t} \quad \dots \quad (3.13)$$

or,
$$\sigma_2^2(t) = \frac{\alpha}{2} (\ln K_0)^2 + \frac{b^2}{4\alpha} + \frac{D}{\alpha} + \frac{2\alpha^2 b \ln K_0}{4\alpha^2 + \omega^2} (2\alpha \cos \omega t + \omega \sin \omega t) +$$

$$+ \frac{b^2 (\alpha \cos 2\omega t + \omega \sin 2\omega t)}{4(\alpha^2 + \omega^2)} - \left\{ \frac{\alpha}{2} (\ln K_0)^2 + \frac{b^2}{4\alpha} + \frac{D}{4\alpha} + \right.$$

$$\left. + \frac{4\alpha^3 b \ln K_0}{4\alpha^2 + \omega^2} + \frac{\alpha b^2}{4(\alpha^2 + \omega^2)} - \sigma^2(0)e^{-2\alpha t} \right\} \quad \dots \quad \dots \quad (3.14)$$

Which gives the pattern of the fluctuation $x(t) = \ln N(t)$. For large time t , $\sigma_2^2(t)$ is finite and oscillatory of superposition of a number of waves, having different amplitudes and frequencies. Assume $\alpha = 1, x_0 = 10, T = 12, \omega = 0.524, K_0 = 100, b = 0.05$,

$D = 1, \sigma^2(0) = 2$, then the Fig. 4 (changes of variance with time in days for periodically varying carrying capacity) represents the graph of time versus variance.

CONCLUSION

Gompertzian Model of growth plays significant role in the mathematical modelling of different biological growth phenomena such as tumour growth [3], growth of plants [2], animals [12]. In recent years non-equilibrium thermodynamic modelling of biological growth is in progress [13,14]. Non-equilibrium thermodynamic theory of biological growth on the basis of Gompertz model equation was developed by Chakrabarti and Bhadra [11], Sarkar and Bhaumik [10].

In this paper we have investigated the characteristic stochastic properties of the Gompertzian growth model for two cases. First one is cyclically growth coefficient having constant carrying capacity with time. Second one is periodically varying carrying capacity having constant growth coefficient with time. In both cases changes of averages and variances (fluctuations) with time have been analysed and the patterns have been shown graphically. Since $x(t)$ is logarithmic and monotonically increasing function of the population size the conclusion drawn for $x(t)$ also holds good for the population size $N(t)$.

ACKNOWLEDGEMENT

The author is thankful to Prof. C.G. Chakrabarti, S.N. Bose Professor of Applied Mathematics, Calcutta University for the help in the preparation of this paper. The author also wish to thank Sri Malay Banerjee, Scottish Church College, Kolkata for the preparation of Computer Graph.

REFERENCES

1. R.B. Banks, Growth and Diffusion Phenomena : Spring-Verlag (1994), Berlin.
2. J. France and J.H.M. Thornley : Mathematical Models in Agriculture, Butterworths, London, (1984).
3. A.K. Laird : Dynamics of Embryonic Growth. 30, (1966), 263-275.
4. T. Malthus : An Essay of the Principle of Population, Reprinted 1970, Penguin Books, Harmondsworth, England, (1798).
5. R.M. Nisbet and W.S.C. Gurney : Modelling Fluctuating Populations. John Wiley & Sons Ltd., New York, (1982).
6. V.S. Pugachev : Theory of Random Functions with Applications, Pergamon Press, Oxford (1963).
7. R.N. Talapatra and C.G. Chakrabarti : St. Cere, Fiz. Tom. 42 Nr. 5 (1990), 421, Bucaresti.
8. C.W. Gardiner : Hand Book of Stochastic Processes. Springer-Verlag, Berlin (1983).
9. B. Gompertz : On the nature of function expressive of the law of mortality. Phil. Trans. 27, (1825), 513.
10. N.G. Sarkar & M.S. Bhaumik : Gompertz Equation : Thermodynamic Modelling of Biological Growth. J. Nat. Phys. Sc. 14 (1-2), (2000), 117-124.
11. C.G. Chakrabarti, & S. Bhadra : Nonequilibrium thermodynamics and stochastics of Gompertz Growth J. Biol. System, 4, (1996), 151.
12. D.M. Euston : Gompertz survival kinetics : fail in number alive or growth in number dead? Theoret. Population Biology, 48, (1995), 1.
13. I. Lamprecht : Application of the concepts of classical Thermodynamics in Biology. In "Thermodynamics of Biological Processes" Eds. I. Lamprecht, & A.I. Zotin Walter de Gruyter, Berlin (1978).
14. R. Walter, & I. Lamprecht : Modern Theories concerning the growth equation. In "Thermodynamics of Biological Processes" Eds. I. Lamprecht, & A.I. Zotin, Walter de Gruyter, Berlin (1978).

THREE FIXED POINT RESULTS FOR THREE MAPS

K.P.R. RAO*, N. SRINIVASA RAO* AND B.V.S.N. HARI PRASAD*

(Received 22.08.2003)

ABSTRACT

We obtain a fixed point Theorem in three different spaces and two more fixed point Theorems in a space using three different non-negative functions.

Key Words : Fixed point, complete metric space, commutativity, non-negative Function, contractive conditions.

2000 Mathematics subject classification : 47H10, 54H25.

INTRODUCTION

We obtain three fixed point theorems using three non-negative functions for three maps satisfying Nung [1], Jain et al. [2] type Contractive conditions.

Nung [1] proved

THEOREM 1 :

Let (X, d) , (X, d) , (Y, ρ) and (Z, σ) be complete metric spaces and

$T: X \rightarrow Y, S: Y \rightarrow Z, R: Z \rightarrow X$ be continuous maps satisfying

$$d(RSTx, RSY) \leq c \max \{d(x, RSy), d(x, RSTx), \rho(y, Tx), \sigma(Sy, STx)\} \quad (1.1)$$

$$\rho(TRSy, TRz) \geq c \max \{\rho(y, TRz), \rho(y, TRSy), \sigma(z, Sy)\}, d(Rz, RSy) \quad (1.2)$$

$$\sigma(STRz, STx) \leq c \max \{\sigma(z, STx), \sigma(z, STRz), d(x, Rz), \rho(Tx, TRz)\} \quad (1.3)$$

for all $x \in X, y \in Y$ and $z \in Z$ where $0 \leq c < 1$. Then RST has a unique fixed point u in X , TRS has a unique fixed point v in Y and STR has a unique fixed point w in Z . Further $Tu = v$, $Sv = w$ and $Rw = u$.

Jain, et al. [2] proved

THEOREM 2 :

Let (X, d) , (Y, ρ) and (Z, σ) be complete metric spaces and

$T: X \rightarrow Y, S: Y \rightarrow Z, R: Z \rightarrow X$ be maps satisfying

* Department of Applied Mathematics, N.U.P.G. Centre, NUZVID A.P. Pin-521 201

$$d^2(RSy, RSTx) \leq c \max \{d(x, RSy)\rho(y, Tx), \rho(y, Tx)d(x, RSTx),$$

$$d(x, RSTx)\sigma(Sy, STx), \sigma(Sy, STx)d(x, RSy)\} \quad (2.1)$$

$$\rho^2(TRz, TRSy) \leq c \max \{\rho(y, TRz)\sigma(z, Sy), \sigma(z, Sy)\rho(y, TRSy)$$

$$\rho(y, TRSy)d(Rz, RSy), d(Rz, RSy)\rho(y, TRz)\} \quad (2.2)$$

$$\sigma^2(STx, STRz) \leq c \max \{\sigma(z, STx)d(x, Rz), d(x, Rz)\sigma(z, STRz),$$

$$\sigma(z, STRz)\rho(Tx, TRz), \rho(Tx, TRz)\sigma(z, STx)\} \quad (2.3)$$

for all $x \in X, y \in Y$ and $z \in Z$ where $0 \leq c < 1$. If one of R, S, T is continuous then RST has a unique fixed point u in X , TRS has a unique fixed point v in Y and STR has a unique fixed point w in Z . Further $Tu = v$, $Sv = w$ and $Rw = u$.

MAIN RESULTS

First we prove a fixed point Theorem in three different spaces.

THEOREM 3 :

Let X, Y, Z be three non-empty spaces, ' d ', ' ρ ', ' σ ' be three non-negative functions from $X \times X$, $Y \times Y$ and $Z \times Z$ such that $d(x_1, x_2) = 0$ iff $x_1 = x_2$, $\rho(y_1, y_2) = 0$ iff $y_1 = y_2$ and $\sigma(z_1, z_2) = 0$ iff $z_1 = z_2$. Further assume that $T: X \rightarrow Y, S: Y \rightarrow Z, R: Z \rightarrow X$ be satisfying

$$d(RSy, RSTx) < \max \{d(x, RSy), d(x, RSTx), \rho(y, Tx)\} \quad (3.1)$$

$$\rho(TRz, TRSy) < \max \{\rho(y, TRz), \rho(y, TRSy), \sigma(z, Sy)\} \quad (3.2)$$

$$\sigma(STx, STRz) < \max \{\sigma(z, STx), \sigma(z, STRz), d(x, Rz)\} \quad (3.3)$$

for all $x \in X, y \in Y, z \in Z$ with $y \neq Sy, x \neq Rz$ respectively.

If the function $x \rightarrow \rho(y, TRSy)$ attains its infimum on Y or the function

$z \rightarrow \sigma(z, STRz)$ attains its infimum on Z then RST or TRS or STR has a unique fixed point.

PROOF :

Suppose the function $x \rightarrow d(RSTx, x)$ attains its infimum on X .

That is there exists a $u \in X$ such that $d(u, RSTu) = \inf \{d(x, RSTx) / x \in X\}$

$$= f(u) \text{ (say)}$$

Suppose RST , TRS and STR have no fixed points.

From (3.1) we have

$$f(RSTRSTu) = d(RSTRSTu, RSTRSTRSTu)$$

$$< \max \{d(RSTRSTu, RSTRSTu), d(RSTRSTu, RSTRSTRSTu), \rho(TRSTu, TRSTRSTu)\}$$

$$= \rho(TRSTu, TRSTRSTu)$$

From (3.2) we have

$$\rho(TRSTu, TRSTRSTu) < \max \{p(TRSTu, TRSTu), \rho(TRSTu, TRSTRSTu),$$

$$\sigma(STu, STRSTu)\}$$

From (3.3) We have

$$\sigma(STu, STRSTu) < \max \{\sigma(STu, STu), \sigma(STu, STRSTu),$$

$$d(u, RSTu)\}$$

$$= d(u, RSTu) = f(u)$$

Thus $f(RSTRSTu) < f(u)$, which is a contradiction.

Hence either RST or TRS or STR has a fixed point.

Similarly we can prove either RST or TRS or STR has a fixed point

when either $y \rightarrow \rho(y, TRSy)$ or $z \rightarrow \sigma(z, STRz)$ attains infimum.

Suppose RST has two fixed points say u, u' .

$$\begin{aligned} d(u, u') &= d(RSTu, RSTu') < \max \{d(u', u), d(u', u'), \rho(Tu, Tu')\} \\ &= \rho(Tu, Tu') \end{aligned}$$

$$\begin{aligned} \rho(Tu, Tu') &= \rho(TRSTu, TRSTu') \\ &< \max \{\rho(Tu', Tu), \rho(Tu', Tu'), \sigma(STu, STu')\} \\ &= \sigma(STu, STu') \end{aligned}$$

$$\begin{aligned} \sigma(STu, STu') &= d(STRSTu, STRSTu') \\ &< \max \{\sigma(STu', STu), \sigma(STu', STu'), d(u, u')\} \\ &= d(u, u') \end{aligned}$$

Then $d(u, u') < d(u, u')$, a contradiction.

Thus RST has a unique fixed point.

REMARK 4 :

Theorem 3 holds if (3.1), (3.2), (3.3) are replaced by

$$\begin{aligned} d^2(RSy, RSTx) &< d(x, RSy)d(x, RSTx) + d(x, RSTx)\rho(y, Tx) \\ &\quad + d(x, RSY)\sigma(Sy, STx) + d(x, RSy)\rho(y, Tx) \end{aligned} \quad (4.1)$$

$$\begin{aligned} \rho^2(TRz, TRSy) &< \rho(y, TRz)\rho(y, TRSy) + \rho(y, TRSy)\sigma(z, Sy) \\ &\quad + \rho(y, TRz)d(Rz, RSy) + \rho(y, TRz)\sigma(z, Sy) \end{aligned} \quad (4.2)$$

$$\begin{aligned} \sigma^2(STx, STRz) &< \sigma(z, STx)\sigma(z, STRz) + \sigma(z, STRz)d(x, Rz) \\ &\quad + \sigma(z, STx)\rho(Tx, TRz) + \sigma(z, STx)d(x, Rz) \end{aligned} \quad (4.3)$$

REMARK 5 :

Theorem 3 holds if (3.1), (3.2) and (3.3) are replaced by

$$\begin{aligned}
 d^3(RSy, RSTx) < \max \{ & d(x, RSy)d^2(x, RSTx) + d^2(x, RSy)d(x, RSTx) \\
 & + d(x, RSy)\rho^2(y, Tx) + d^2(x, RSy)\rho(y, Tx) \\
 & + d(x, RSy)\sigma^2(Sy, STx) + d^2(x, RSy)\sigma(Sy, STx) \\
 & + d(x, RSy)d(x, RSTx)\rho(y, Tx) \\
 & + d(x, RSy)\rho(y, Tx)\sigma(Sy, STx) \\
 & + d(x, RSy)d(x, RSTx)\sigma(Sy, STx), \\
 & d^2(x, RSTx)\rho(y, Tx), d(x, RSTx)\rho^2(y, Tx) \} \quad (5.1)
 \end{aligned}$$

$$\begin{aligned}
 \rho^3(TRz, TRSy) < \max \{ & \rho(y, TRz)\rho^2(y, TRSy) + \rho^2(y, TRz)\rho(y, TRSy) \\
 & + \rho(y, TRz)\sigma^2(z, Sy) + \rho^2(y, TRz)\sigma(z, Sy) \\
 & + \rho(y, TRz)d^2(Rz, RSy) + \rho^2(y, TRz)d(Rz, RSy) \\
 & + \rho(y, TRz)\rho(y, TRSy)\sigma(z, Sy) \\
 & + \rho(y, TRz)d(Rz, RSy)\sigma(z, Sy) \\
 & + \rho(y, TRz)\rho(y, TRSy)d(Rz, RSy), \\
 & \rho^2(y, TRSy)\sigma(z, Sy), \rho(y, TRSy)\sigma^2(z, Sy) \} \quad (5.2)
 \end{aligned}$$

$$\begin{aligned}
 \sigma^3(STx, STRz) < \max \{ & \sigma(z, STx)\sigma^2(z, STRz) + \sigma^2(z, STx)\sigma(z, STRz) \\
 & + \sigma(z, STx)d^2(x, Rz) + \sigma^2(z, STx)d(x, Rz) \\
 & + \sigma(z, STx)\rho^2(Tx, TRz) + \sigma^2(z, STx)\rho(Tx, TRz) \}
 \end{aligned}$$

$$\begin{aligned}
 & +\sigma(z, STx)\sigma(z, STRz)d(x, Rz) \\
 & +\sigma(z, STx)d(x, Rz)\rho(Tx, TRz) \\
 & +\sigma(z, STx)\sigma(z, STRz)\rho(Tx, TRz), \\
 & \sigma^2(z, STRz)d(x, Rz), \sigma(z, STRz)d^2(z, Rz) \}
 \end{aligned} \tag{5.3}$$

Now we give a Theorem on a set equipped with three different non-negative functions.

THEOREM 6 :

Let X be a non-empty set, $S, T, R: X \rightarrow X$ and d, ρ, σ be three non-negative functions from $X \times X$ such that $d(x_1, x_2) = 0$ iff $x_1 = x_2$, $\rho(x_1, x_2) = 0$ iff $x_1 = x_2$ and $\sigma(x_1, x_2) = 0$ iff $x_1 = x_2$. Further assume that

$$d(STx, TRy) < \max \{ \rho(Tx, Ry), d(STx, Tx), \rho(TRy, Ry), \sigma(Tx, TRy) \} \tag{6.1}$$

$$\rho(TRx, RSy) < \max \{ \sigma(Rx, Sy), \rho(TRx, Rx), \sigma(RSy, Sy), d(Rx, RSy) \} \tag{6.2}$$

$$\sigma(RSy, STx) < \max \{ d(Sy, Tx), \sigma(RSy, Sy), d(STx, Tx), \rho(Sy, STx) \} \tag{6.3}$$

for all $x, y \in X$ with $Tx \neq Ry$, $Rx \neq Sy$, $Sy \neq Tx$ respectively.

If the function $x \rightarrow d(STx, Tx)$ attains its infimum on X or the function $x \rightarrow \rho(TRx, Rx)$ attains its infimum on X or the function $x \rightarrow \sigma(RSx, Sx)$ attains its infimum on X then S or T or R has a fixed point.

PROOF :

Assume that $x \rightarrow d(STx, Tx)$ attains its infimum on X . Then there exists $z \in X$ such that $d(STz, Tz) = \inf \{ d(STx, Tx) / x \in X \}$

$$= f(z) \quad (\text{say})$$

Suppose R, S and T have no fixed points in X .

From (6.1) we have

$$\begin{aligned}
 f(RSTz) &= d(STRSTz, TRSTz) \\
 &< \max\{\rho(TRSTz, RSTz), d(STRSTz, TRSTz), \rho(TRSTz, RSTz), \\
 &\quad \sigma(TRSTz, TRSTz)\} \\
 &= \rho(TRSTz, RSTz) \quad \text{since } T \text{ has no fixed point.}
 \end{aligned}$$

From (6.2) we have

$$\begin{aligned}
 \rho(TRSTz, RSTz) &< \max \{\sigma(RSTz, STz), \rho(TRSTz, RSTz), \sigma(RSTz, STz), \\
 &\quad d(RSTz, RSTz)\} \\
 &= \sigma(RSTz, STz) \quad \text{since } R \text{ has no fixed point.}
 \end{aligned}$$

From (6.3) we have

$$\begin{aligned}
 \sigma(RSTz, STz) &< \max \{d(STz, Tz), \sigma(RSTz, STz), d(STz, Tz), \rho(STz, STz)\} \\
 &= d(STz, Tz) \quad \text{since } S \text{ has no fixed point.} \\
 &= f(z)
 \end{aligned}$$

Thus we have $f(RSTz) < f(z)$, a contradiction.

Hence R or S or T has a fixed point.

Finally we prove the following Theorem for commuting maps.

THEOREM 7 :

Let X be a non-empty set, ' d ', ' ρ ', ' σ ' be non-negative functions on $X \times X$ such that $d(x_1, x_2) = 0$ iff $x_1 = x_2$, $\rho(x_1 x_2) = 0$ iff $x_1 = x_2$ and $\sigma(x_1 x_2) = 0$ if $x_1 = x_2 = 0$ iff $x_1 = x_2$. Let $R, S, T: X \rightarrow X$ be commuting pair wise such that

$$d(RSy, RSTx) < \rho(y, Tx) \text{ for } y \neq Tx, RSy \neq RSTx \quad (7.1)$$

$$\rho(TRz, TRSy) < \sigma(z, Sy) \text{ for } z \neq Sy, TRz \neq TRSy \quad (7.2)$$

$$\sigma(STx, STRz) < d(x, Rz) \text{ for } x \neq Rz, STx \neq STRz. \quad (7.3)$$

Then ' R ' has a fixed point if the function $x \rightarrow d(x, Rx)$ attains its infimum on X ,

'S' has a fixed point if the function $x \rightarrow \sigma(x, Sx)$ attains its infimum on Y and 'T' has a fixed point if the function $x \rightarrow \rho(x, Tx)$ attains its infimum on Z .

PROOF :

Assume the function $x \rightarrow d(x, Rx)$ attains infimum at $z \in X$.

That is $d(z, Rz) = \inf \{d(x, Rx) / x \in X\} = f(z)$ (say.)

Suppose 'R' has no fixed point. Then

$$f(RSTRSTz) = d(RSTRSTz, RRSTRSTz)$$

$$= d(RSTRSTz, RSTRSTRz) \quad (\text{since } R, S, T \text{ are commuting})$$

in pairs) $< \rho(TRSTz, TRSTRz)$ from (7.1) since $R(TRSTz) \neq TRSTz$

$< \sigma(STz, STRz)$ from (7.2) since $R(STz) \neq STz$.

$< d(z, Rz)$ from (7.3) since $Rz \neq z$.

$= f(z)$, a contradiction.

Therefore R has a fixed point.

Similarly we can prove the remaining cases.

ACKNOWLEDGEMENTS

The authors wish to thank Prof. I.H.N. Rao and Prof. S.V.R. Naidu for their valuable help in the preparation of the paper.

REFERENCES

1. N. P. Nung : A fixed point Theorem in three metric spaces, Math. Sem. Notes, kobe Univ, 11 (1983), 77-79.
2. R. K. Jain, A. K. Shrivastava, B. Fisher : Fixed points on three complete metric spaces, Novi SAD J. Math Vol. No.1 (1997), 27-35.

ANTIFUNGAL ACTIVITY OF CERTAIN PLANTS AGAINST *ALTERNARIA BRASSICAE AND ASPERGILLUS NIGER*

R.D. KAUSHIK, GARIMA SHARMA & CHARU ARORA

(Received 03.01.03 and in revised form 22.08.03)

ABSTRACT

The present paper deals with the degree of antifungal activity of four types of aqueous extracts of 30 plant species from Kumaon, Garhwal and Tarai regions against two fungal organisms, *Alternaria brassicace* and *Aspergillus niger*. Results revealed that aqueous extracts of eight plants were effective against *A. brassicace* alone but only two of them, *Boenninghausenia albiflora* and *Cannabis sativa*, were found effective against *A. niger* alone. The extracts of five plants, *Aegle marmelos*, *B. albiflora*, *Cleome viscosa*, *Leonotis nepetaefolia*, and *Vitex negundo* have been found to show antifungal activity against both of the fungi. However, the aqueous extracts of fifteen plants could show a low degree of antifungal activity. The specific and common inhibitory activities of plant extracts against these fungi might be interpreted due to presence of selective or wide-spectrum chemical compounds present in the plants.

Keywords: Thirty plants, Aqueous fresh/dry and cold/hot extracts, Antifungal activity, *Alternaria brassicace* and *Aspergillus niger*.

INTRODUCTION

Albeit a large number of plant diseases have been controlled by using synthetic chemicals such as fungicides, pesticides and herbicides since long time, still it is in practice. Voluminous informations are available in this regard and some of the important contributions can be quoted [3, 5-6, 9, 19, 29]. Due to adverse effect of these chemicals in environment and their hard resistance for biodegradation, researchers have opened a new application of plants as biological agents for controlling the pests and plant pathogens that has taken an important place in pest and plant diseases' management programme since last few decades in India and abroad because they do not have such adverse effect on biological and physical system of our environment. These may contain many active ingredients although in low concentration. Degree of antifungal, antibacterial and antiviral property depends on nature and amount of the chemical substances produced by the plant system and stored form time to time in the different plant parts and their solubility at the time of their application. The pests and pathogens do not develop resistance to them, because it might require several simultaneous mutations to occur in the genetic constituents

of the pests and pathogens in order to overcome the ingredients of the botanical pesticides [6, 24]. Plants are used for control of various bacterial, fungal and viral diseases of plants [6, 8, 10-12, 20-21, 24-25]. The chemical characterization of many of these plants has already been done [7, 11, 16, 26-27]. Systematic studies on the role of biocidal compounds from these plants for crop protection have not yet been made due to the lack of proper screening formats and indicator organisms of the plant diseases.

In the present study, 35 plant spp. suspected to possess bioactive chemicals were selected on the basis of information available in literature [7, 18, 20, 26-27], local folklore and field observations on plants that remain relatively free from the diseases. They are present at different height and locations in Kumaon, Garhwal, and Tarai region. However, only 30 plant spp. could be located and used for screening their fungicidal potential against the fungi, *A. brassicaceae* and *A. niger* which cause the leaf spot disease in brassicae and aflatoxin toxicity in maize and groundnut respectively. Though some of these plants have already been reported to show fungicidal action against *A. brassicaceae* and *A. niger* in the form of methanol extracts [4], they have not been screened for their activity in the form of aqueous extracts against the same fungi [18, 20, 27]. Many of these plants have been explored for their antibacterial activity recently [17].

MATERIALS AND METHODS

Collection of plants: All 30 plant spp. included in the present investigations have already been reported to exhibit medicinal / pesticidal or other economic values. These were collected from various places in Kumaon, Garhwal and Tarai region at latitude range $29^{\circ} 01' 20.0''$ to $29^{\circ} 47' 52.1''$, longitude range $79^{\circ} 26' 41.9''$ to $79^{\circ} 35' 05.8''$ and altitude range 228 m to 2230 m. Different parts of the plants used for investigations are given in the table-1. The collected material was carried in polythene bags to the laboratory and dried in shade for 22 days.

Collection of fungal cultures: Fungal cultures of *A. brassicaceae* and *A. niger* were isolated from the hosts, brassicae and maize and tested for Kotch postulate at Department of Plant pathology, College of Agriculture, G.B. Pant university of Agriculture and Technology, Pantnagar, India from where these were procured by us. They were cultured in our laboratory in petriplates containing Patato Dextrose Ager (PDA) media.

Preparation of plant extract: 10 g of plant material collected from natural habitat was extracted four times with 2 ml. solvent/g of plant material at room temperature for preparation of fresh cold water extract (FCW) and in boiling conditions (temperature about 100°C) for the fresh hot water extract (FHW). Remaining plant material was dried on shade. Dry powder of leaves/other parts of plant spp. was extracted four times with 5 ml. solvent/g of plant material for 48 hrs at room temperature or 100°C for preparation of DCW and DHW respectively. All these extracts were combined and concentrated by flash evaporation at 40°C to 2 ml (i.e. 20 µl extract equivalent to 100 mg of plant material). five factors were involved in the treatment i.e. (a) Fresh cold water extract (FCW) (b) Fresh hot water extract (FHW) (c) Dry cold water extract (DHW) (d) Dry hot water extract (DHW) and (e) two fungal pathogens, *A. brassicae* and *A. niger*.

Screening of extracts for antifungal activity: Paper disc method was used for screening anti-fungal activity of plant extract against test fungi. This method was based on diffusion capacity of test chemical (s) through an agar medium. Fungal plugs were placed at the centre of PDA plates and allowed to grow. After circular growth of about 2-3 cm diameter, four sterilized paper discs (two loaded with 20 µl plant extracts and two with same amount of pure solvent) were placed at equal distance from center in order to see the antifungal potential of test plants against test fungi. These plates were incubated at 28 ± 1°C for both the fungal pathogens. Inhibition zones were measured after four days of incubation. Dumb bell shaped growth of fungi was observed in case of extracts containing bioactive components. All the tests were carried out in triplicate and the mean values were used for interpretation.

RESULTS AND DISCUSSION

The extracts of *Acorus calamus* L., *Artemisia nilagirica* (Clarke) Pamp., *Cassia fistula* L., *Cassia occidentalis* L., *Ficus religiosa* L., *Hedychium spicatum* Smith., *Iris Kumaonesis* Wall., *Justicia adhatoda* L., *Lantana camara* L., *Litsea glutinosa* (Lour.) Robins., *Polygonum lapathifolium* L., *Sapium insigne* (Royle) Benth. ex. Hook. f., *Smilax aspera* L., *Urtica dioica* L. and *Viola biflora* L. were found ineffective against both of the fungal pathogens.

The results of fungicidal activity of other fifteen plants are presented in the table-1. *A. niger* is more resistant towards all plant extracts except those of *Cannabis sativa* L. and *Toona ciliata* Roem., the former being more potent (inhibition zone of

2-3 mm) in comparison to the latter (inhibition size 1mm. only). *Berberis aristata* DC., *Conyza bonariensis* (L.) Crong., *Erigeron karvinskianus* DC., *Lepidium apetalum* Willd., *Lyonia ovalifolia* (Wall.) Drude., *Mentha longifolia* (L.) Hudson., *Thalictrum foliolosum* DC., and *Woodfordia fruticosa* (L.) Kurz. are potent against *A. brassicae* only. While all of them show inhibition zone between 1-3 mm., *W. fruticosa* is most effective with zone size of 5-6 mm. The plant extracts of *Aegle marmelos* (L.) Corr., *Boenninghausenia albiflora* (Hook.) Reichb., *Cleome viscosa* L., *Leonotis nepetaefolia* (L.) W. Ait., and *Vitex negundo* L. are effective against both of the fungi involved in the present investigation with inhibition between 1-6 mm. However, *C. viscosa* and *V. negundo* extracts are better in fungidal potential against *A. brassicae* while other plant spp. are more potent against both of the test fungi.

A perusal of the data indicates that the solubility of the active constituents of extracts of some of the plant spp. are not influenced and no degradation takes place on heating up to 100°C while in some cases, these parameters are expected to be enhanced or suppressed resulting in the increased or decreased fungicidal action respectively.

The reduced activity (when the hot extracts are used in some of the cases), may be due to the decreased solubility and degradation/deterioration of the active fungicidal ingredients of the plants used. Basic principles of the chemical energetics do not rule out the chances of reduction of solubility of certain chemical substances on elevating the temperature when the molar enthalpy of solubilization bears a negative value. Similarly, the increased antifungal activity on using hot aqueous extracts may be due to the increased solubility and activation of the active principles of plants involved. Increase in the fungicidal potential in some of the plants on using DHW in place of FHW implies the better exposure of the active components on drying of plant material.

The fungicidal action of the extracts of *V. negundo*, *C. viscosa* and *W. fruticosa* against *A. brassicae* is adversely affected on using the hot extracts while it is the reverse case or no change is observed against *A. niger*. It points towards the possibility that different active principles of the same plant are acting as fungicidal agents against *A. brassicae* and *A. niger*. Actually, these plant spp. have shown the best activity against *A. brassicae* amongst all the 30 plants used for this study.

It is encouraging to find that the ornamental plants like *C. karvinskianus* and

E. bonariensis [13] can be used as fungicidal agents. Further, these are far better in fungicidal potential especially against *A. brassicae* in comparison to the plants *A. calamus*, *L. ovalifolia* and *T. ciliata* which are better known for their fungicidal other fungi [6, 2, 24] or insecticidal plants like *C. occidentalis*, *L. glutinosa* and *S. insigne* [8, 27] or bactericidal plant like *L. nepetaefolia* [10] or *P. lapathifolium*- a pesticidal plant [25]. Their fungicidal action is better than the medicinal plants like *A. marmelos*, *A. nilagirica*, *B. albiflora*, *C. fistula*, *F. religiosa*, *J. adhatoda*, *T. foliolosum*, *U. dioica* and *V. biflora* [18, 22-23, 26-28] which, in turn, are either not effective at all or have comparatively lesser fungicidal potential against the fungi used by us.

Many of these plants have been reported to exhibit antifungal activity when their aqueous/ methanot extracts were used against the fungal pathogens of rice like *Magnaporthe grisea* and *Rhizoctonia solani* [14-15] or against mushroom fungal pathogens like *Mycogone perniciosa* and *Verticillium fungicola* [1] or against the soybean fungal pathogens like *Colletotrichum truncatum*, *Fusarium oxysporum* and *Macrophomina phaseolina* [2].

We have recently reported the fungicidal potential of the methanol extracts of dry powder of the leaves of these and some other plants against *A. brassicae* and *A. niger* [4]. A comparison with the reported work [4] indicates that the degree of antifungal activity is not as good in aqueous extracts (both DCW and DHW) as in the case when the dry methanol extracts are applied. However, it is worth mentioning that water (and not the methanol), is readily available as a solvent in the natural system. Therefore, the active antifungal principles of plants are expected to be used in the form of aqueous extracts in the natural ecosystems. It makes the aqueous extracts more significant as the careers for the natural fungicides.

The effective plant extracts are expected to contain wide spectrum antifungal chemical (s). These plant species may contain chemical compound (s) having selective/ specific antifungal properties. Therefore, the results described above may be helpful in developing/ synthesizing the plant based natural fungicides that may be used for preventing the incidence of leaf spot of brassicae and aflatoxin toxicity in maize, groundnut etc.

ACKNOWLEDGEMENT

The authors are thankful to Dr. Anil Kumar and Dr. Suresh Singh, G.B. Pant University of Agriculture and Technology, Pantnagar (India) for their help in collection

Table-1 : Inhibitory Effect of aqueous extracts of plants against *A. brassicace* and *A. niger*.

S. No.	Plant species (Part used)	Type * of Extract	Zone size (mm)		S. No.	Plant species (Part used)	Type of Extract	Zone size (mm)	
			<i>A. brassicace</i>	<i>A. niger</i>				<i>A. brassicace</i>	<i>A. niger</i>
1.	<i>A. marmelos</i> (Leaf)	FCW	2	0	9.	<i>L. apetalum</i> (Leaf)	FCW	0	0
		FHW	2	4			FHW	0	0
		DCW	3	2			DCW	1	0
		DHW	2	6			DHW	2	0
2.	<i>B. aristata</i> (Leaf)	FCW	2	0	10.	<i>L. ovalifolia</i> (Leaf)	FCW	1	0
		FHW	2	0			DCW	2	0
		DCW	3	0			DCW	1	0
		DHW	3	0			DHW	1	0
3.	<i>B. albiflora</i> (Whole plant)	FCW	0	1	12.	<i>T. foliolosum</i> (Bark)	FCW	1	0
		FHW	0	1			FHW	2	0
4.	<i>C. sativa</i> (Leaf)	FCW	0	2			DCW	2	0
		DHW	0	3			DHW	2	0
5.	<i>C. bonariensis</i> (Leaf)	FCW	1	0	13.	<i>T. ciliata</i> (Bark)	FCW	0	1
		FHW	0	0			FHW	0	1
		DCW	2	0			FCW	5	2
		DHW	2	0			FHW	4	3
6.	<i>C. viscosa</i> (Leaf)	FCW	2	2	14.	<i>V. negundo</i> (Leaf)	DCW	6	4
		FHW	1	0			DHW	5	4
		DCW	6	5			FCW	6	0
		DHW	6	5			FHW	5	0
7.	<i>E. karvinskianum</i> (Leaf)	DCW	3	0	15.	<i>W. fruticosa</i> (Leaf)	DCW	6	4
		DHW	2	0			DHW	5	4
		FCW	0	2			FCW	6	0
		FHW	0	2			FHW	5	0
8.	<i>L. nepetaefolia</i> (Leaf)	DCW	1	2					
		DHW	1	3					
		FCW	0	2					
		FHW	0	2					

*FCW – Fresh cold aqueous extract ; FHW – Fresh hot aqueous extract
 DCW – Dry cold aqueous extract ; DHW – Dry hot aqueous extract

of the fungal cultures and in identification / collection of plant species respectively.

REFERENCES

- Charu Arora, R. D. Kaushik, A. Kumar and G. K. Garg : Fungicidal potential of Kumaon and Tarai plants against mushroom fungal pathogens, Allelopathy journal, 11 (1) (2003), 63-70.

2. Charu Arora and R.D. Kaushik: Fungicidal activity of plants extracts from Uttarakhand hills against soybean fungal pathogens, *Allelopathy Journal*, 11(2) (2003), 217-228.
3. Charu Arora and R.D. Kaushik : Antifungal activity of some transition metal ferrocyanides, *Asian J.Chem.*, 15(3-4) (2003), 1828-1830.
4. Charu Arora and R.D. Kaushik : Fungicidal potential of some plants from Uttarakhand hills against *A. brassicae* and *A. niger*, *J. Curr. Sci.*, 3(1) (2003), 47-52.
5. R. Cremlyn: Pesticides-Preparation and mode of action. John Wiley & Sons., (1978), 141-142.
6. M. Damayanti, K. Susheela and G. J. Sharma : Effect of plant extracts and systemic fungicide on the pineapple fruit-rotting fungus *Ceratocystis paradoxa*, *Cytobios*, 86 (1996), 155-165.
7. B. Das and R. Das : Medicinal properties and chemical constituents of *Vitex negundo* L., *Indian Drugs*, 31 (9) (1994), 431-435.
8. S. Facknath and D. Kawol : Antifeedant and insecticidal effects of some plant extracts on the cabbage webworm, *Crocidolomia binotalis*, *Insect Science and its Application*, 14 (5) (1993), 571-574.
9. F.J. Gee, J.C. Tello and M. Honrubia : In vitro sensitivity of *Verticillium fungicola* to selected fungicides, *Mycopathologia*, 136(3) (1997), 133-137.
10. R. H. Gopal, V. Saradha, K.E. Vinnarasi, S. Govindarajan and S. Vasanth : Antibacterial activity of *Leonotis nepetaefolia*, *Fitoterapia*, 66 (1) (1995), 83-84.
11. A. Haraguchi, R. Matsuda and K. Hashimoto: High performance liquid chromatographic determination of sesquiterpene dialdehydes and antifungal activity from *Polygonum hydropiper*, *J. Agric. Food Chem.*, 41(1993), 5-7.
12. R. Jagannathan and K. Sivaprakasan: Effect of botanicals on managing sheath rot of rice, *International Rice Research Notes*. 21 (1) (1996), 49-50.
13. R. Kadner: Bedding and balcony plants in hanging baskets, *Gartenbau Megazin.*, 1(5) (1992), 70.
14. R. D. Kaushik and Charu Arora: Antifungal activity of aqueous extracts of some economically important plants against *M. grisea* and *R. solani*, *Indian J. Environ. & Ecopl.*, 7(2) (2003), 273-278.
15. R. D. Kaushik and Charu Arora : Fungitoxic activity of methanol extracts of some plants of Kumaon, Garhwal and Tarai regions against fungal pathogens of rice, *J. Indian Bot. Soc.*, 81(3-4)(2002), 327-331.
16. R. D. Kaushik, Anil Kumar and Charu Arora : Fractionation, Characterization and evaluation of biocidal potential of active principles of leaves of *Vitex negundo* Linn., *Asian J. Chem.*, 15(3-4)(2003), 1659-1664.

17. R. D. Kaushik, G. K. Garg, Garima Sharma and Charu Arora : Antibacterial activity of plant extracts from Uttaranchal Hills, India, Allelopathy Journal, 12(2)(2003), 205-214.
18. K. R. Kirtikar and B.D. Basu: Indian Medicinal plants. M/S Bishen Singh Mahendra Pal Singh, Dehradun (India), Ed. 2, Vol.III, (1980).
19. T. X. Liu, P. A. Stansly and O. T. Chortyk: Insecticidal activity of natural and synthetic sugar esters against *Bermisia argentifolii* (Homoptera: Aleyrodidae), Journal of Economic Entomology, 89 (1996), 1233-1239.
20. S. A. H. Naqvi, M. S. Y. Khan and S. B. Vohra: Antibacterial, antifungal and anthelmintic investigations of Indian medicinal plants, Fitoterapia, 62(3)(1991), 221-228.
21. J.C. Pandey, R. Kumar and R. C. Gupta: Possibility of biological control of rhizome rot ginger by different antagonists, Progressive Horticulture, 24 (3-4)(1992), 227-232.
22. C. Randall, K. Meethan, H. Randall and F. Dobbs : Nettle sting of *Urtica dioica* for joint pain-an exploratory study of this complementary therapy, Therapies in Medicine, 7(3) (1999), 126-131.
23. M.A. Rashid and F. Ahmad : Identification of the natural drug 'Biranjasaf' (*Artemisia nilagarica*)-A comparative pharmacognostical study, Hamdard-Medicus, 39(2)(1996), 46-57.
24. U. Sardsud, V. Sardsud, C. Sittigul, T. Chaiwangsri, G. I. Johnson (ed.) and E. Highley : Effects of plant extracts on the in vitro and in vivo development of fruit pathogens, Workshop on "Development of post harvest handling technology for tropical tree fruits", Bangkok, Thailand, 16-18 July, (1992), 60-62.
25. P. B. Sas, and W. Piotrowski : The growth and development of potato pathogens on the media with extracts from polygonaceae plants. 2. Pathogens causing tuber diseases, Biuletyn Instytutu Ziemniaka, 48 (II)(1997), 91-99.
26. V. K. Singh and Z. A. Ali : Aspects of plant Sciences, Vol. 15, Harbal Drugs of Himalaya (Medicinal plants of Garhwal and Kumaon regions of India), Today and Tomorrow's printers and publishers, N. Delhi, (1998), 1, 220.
27. The useful plants of India, CSIR, New Delhi (India). (1986,) 682.
28. L. Udupa,A. L. Udupa and D. R. Kulkarni : Studies on the anti-inflammatory and wound healing properties of *Moringa oleifera* and *Aegle marmelos*, Fitoterapia, 65(1994), 119-123.
29. A. V. Zaayen : Prochloraz (sporon), a new fungicide in mushroom culture, Champignoncultuur, 2(4)(1983), 163-167.

SOME GEOMETRICAL CONSTRUCTIONS FROM BAUDHĀYANA ŚULBA SŪTRA

V.K. SHARMA* AND YOGITA BANA*

(Received 08.10.2003)

ABSTRACT

Some geometrical constructions from Baudhāyana Śulba Sūtra are given.

Key words and Phrases : Square, Side, Tie, Pole, Rectangle, Oblong, Circle, Triangle, Prauga, Srori, Cord, Prāci and Area etc.

Classification Number : 01A32

INTRODUCTION

Sulba Sūtra-

The science of mathematics which is said to be the “queen of all sciences” or “the peak of all sciences” has originated in so hoary manner that it is impossible to trace its origin in any of the written records of the human civilization. Mathematics was known in ancient India as Ganit-veda or the Veda (i.e. knowledge) of calculations See 6 & 7.

India's oldest written works are the Vedas. There are four Vedas, namely Rgveda (ऋग्येद) which is the oldest, Yajurveda, Sāmaveda, and Atharvaveda also, associated with a Veda. There are different branches or schools which are represented by the various Saṁhitās (or recensions) of that Veda. To assist the proper study of the Veda, there are Six Vedāngas, ("limbs or parts of the Veda"). Namely Śiksā (Phonetics), Kalpa (Ritualistics), Vyākaraṇa (Grammar), Nirukta (Etymology), Chands (Prosody), and Jyotiṣa (Astronomy and Astrology), as is clear from the popular verse See 1,3 & 5.

शिक्षा कल्पो व्याकरणं निरुक्तं छन्दसांचयः ।

ज्योतिषामयनं चैव वेदाङ्गानि षडेव तु ॥

The Kalpa deals with rules and methods for performing vedic rituals, sacrifices and ceremonies, and is divided into three categories which are called Śrauta (श्रौत), Grhya (गृह) and Dharma (धर्म) See 5.

The Śrauta Sūtras, especially those belonging to the various sāmhitās of the Yajurveda. Often include tracts which give rules concerning the mensuration and construction of vedis

*Department of Mathematics & Statistics, Gurukula Kangri Vishwavidyalaya, Haridwar.

(sacrificial grounds), citis (mounds or altars), and agnis (fire-places). Such tracts are also found as separate works and are called Śulba Sūtras (शुल्ब सूत्राणि) (which are denoted by SS here after). Thus SS are the oldest geometrical treatises which represent, in coded form, the much older and traditional Indian mathematics developed for construction and transformation of vedic altars of various types and forms. The word śulba literally means a cord, rope, or string and is derived from the basic root śulb (or śulv) meaning “to mete out” or “to measure” See 5.

By now the names of about a dozen SS are known. They are Baudhāyana, Āpastamba, Kātyāyana, Mānava, Satyāśādha, Maitrāyana, Varāha, Vadhula, Maśaka, Hiranyakeshin, and Laugākgi. They are variously dated their exact times of composition or compilation is controversial. The Baudhāyana Śulba Sūtra (=BSS) which is the oldest of them is generally placed between 800 B.C. and 500 B.C. The Āpastamba Śulba Sūtra (=ASS), Kātyāyana Śulba Sūtra (=KSS), and Mānava Śulba Sūtra (=MSS) are other important ancient works of the class. The text of Satyāśādha is said to be identical with that of ASS while Maitrāyana is said to be another recension of the MSS See 5.

SOME GEOMETRICAL CONSTRUCTIONS

1. TO CONSTRUCT A SQUARE WITH THE HELP OF A CORD ALONE

चतुरस्त्रं चिकीर्षन्यावच्चिकीर्षत्तावीर्जुमुभयतः पाशां कृत्वा मध्ये लक्षणं करोति लेखामालिख्य ॥

(BSS. 1/22)

If you wish to make a square, take a cord of the length which you desire to give to the side of the square, make a tie at both its ends and a mark at its middle; then having drawn the line (i.e., praci, the line pointing to the east and west points; the different methods by which this line was found will be discussed later.)

तस्या मध्ये शङ्कुं निहन्यात्सिम्न्याशौ प्रतिमुच्य लक्षणेन मण्डलं परिलिखेत् विश्वकम्भान्तयोः शङ्कुं निहन्यात् ॥ (BSS. 1/23)

Fix a pole in the middle of that line; having fastened at that pole the two ties of the cord, describe (around the pole) a circle with the mark. (As the mark divides the length of the cord, the circle is to be described with half the length of the cord); fix poles at both ends of the diameter (i.e., diameter formed by the praci).

पूर्वस्मिन्याशं प्रतिमुच्य पाशेन मण्डलं परिलिखेत् ॥ (BSS. 1/24)

Having fastened one tie at the eastern pole (the pole fix at the east and of the praci) described a circle with the other tie (i.e., with the full length of the cord).

एवमपरस्मिन्पुरुष्टे यत्र समेयातां तेन द्वितीयं विष्कम्भमायच्छेत् ॥ (BSS. 1/25)

In the same manner, a circle describe round the pole at the west end of the prāci; then a second diameter is drawn joining the points in which these two circles intersect (this diameter is the line pointing to the north and south points).

विष्कम्भान्तयोः शङ्कू निहन्यात् ॥ (BSS. 1/26)

Fix a pole at both ends of the diameter.

पूर्वस्मिन्याशौ प्रतिमुच्य लक्षणेन मण्डलं परिलिखेत ॥ (BSS. 1/27)

Having fastened both ties at the east pole, describe a circle round it with the mark.

एवं दक्षिणात् एवं पश्चादेवमुत्तरतस्तेषां येऽन्त्यः संसर्गास्तच्यतुरस्त्रे संपद्यते ॥ (BSS. 1/28)

The same is to be done in the south, the west, and the north (i.e., circles are to be described round the three other poles); the points of intersection of these four circles which (i.e., the points) are situated in the four intermediate regions (north-east, north-west etc.) are the four corners of the required square.

2. TO CONSTRUCT A SQUARE EQUAL TO THE SUM OF DIFFERENT SQUARES

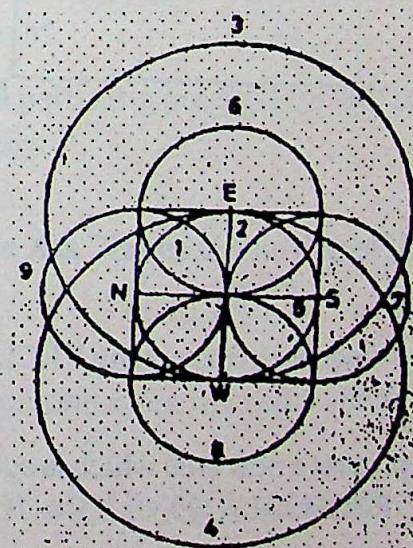
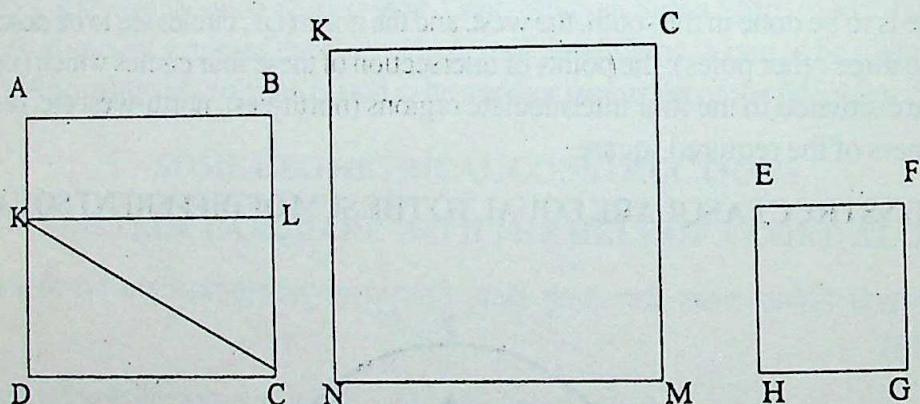


Fig. 1 : (Adhyaya 1, Sutra 28)

नाना चतुरस्त्रे समस्यन्कनीयसः करण्या वर्षीयसो वृध्मुल्लिखेद्वधस्याक्षणायाज्जुः समस्तयोः पाश्वर्मानी भवति ॥ (BSS. 1/50)

If you wish to combine two squares of different size into one, scratch up with the side of the smaller square a piece cut-off from the large one (i.e., cut-off a piece from the large square by scratching up the ground-or making a mark upon the ground-at a distance from one end of a side of the large square, which is equal to the length of the side of the smaller square; by repeating this process on the opposite side of the large square and joining the two marks on the ground by a line or cord, and oblong is cut-off, of which the two longer to sides are equal to the side of the large square and the two shorter sides to the side of the small square). The diagonal of this cut-off piece is the side of the combined squares (of the square which combines the two squares).



ABCD is a big square. Cut EH length on AD and BC at K and L respectively. Join KL and KC.

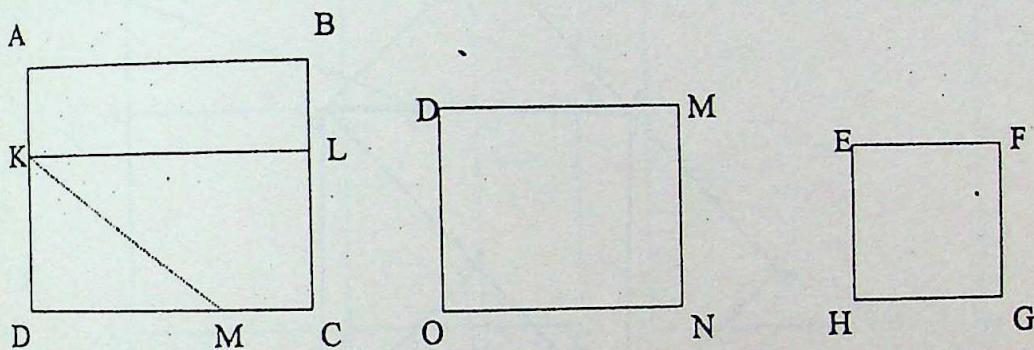
An oblong KLCD is cut off, the diagonal of the oblong KLCD is the side of the combined squares.

Hence square KCMN will represent a new area which is equal to the sum of area of the sq. ABCD and area of the sq. EFGH.

3. TO CONSTRUCT A SQUARE EQUAL TO THE DIFFERENCE OF TWO SQUARES

चतुरस्त्राच्चतुरस्त्रं निर्जिहीर्षन्यावनिनर्जिहीर्षत्तस्य करण्या वर्षीयसो वृध्मुल्लिखेत् । वृधस्य पाश्वर्मानीमद्धणयेतरत्पाश्वर्मुपस्तुहरेत्सा यत्र निपतेतदपच्छिन्द्याच्छिन्नया निरस्तम् ॥ (BSS. 1/51)

If you wish to deduct one square from another square, cut-off a piece from the large square by making a mark on the ground with side of the smaller square which you wish to deduct (an oblong is cut-off, the sides of which are equal to the sides of the two given squares); draw one of the sides (the cord representing one of the longer sides of the oblong) across the oblong so that it touches the other side; where it touches (the other side), by this line which has been cut-off the small square is deducted from the large one (i.e., the cut-off the line is the side of a square the area of which is equal to the difference of two squares.)



ABCD is a big square. Cut EH length on AD and BC at K and L respectively. Join KL cut the length KL at the DC taking form K as the center put an are equal to KL on the base DC. Let we name this point as M.

Now base is divided into two parts. Big portion DM will be side of a required square because in a right angled triangle.

$$KD^2 + DE^2 = KM^2$$

$$EF^2 + DM^2 = AB^2$$

$$\text{or } DM^2 = AB^2 - EF^2$$

Hence square DMNO will represent a new area after subtracting area of the sq. EFGH from sq. ABCD.

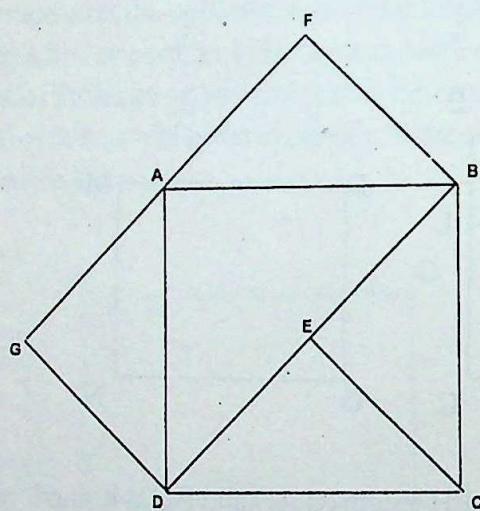
4. TRANSFORMING A SQUARE INTO A RECTANGLE

समचतुरस्रं दीर्घचतुरस्रं चिकीर्षा॒ स्तदक्षायापच्छया॑ भागं द्वेषा॒ विभज्य पाश्वयोरुपदध्याद्यथायोगम् ॥
(BSS.1/52)

To convert a square into rectangle Baudhāyana suggests that divide a square into

two parts by a diagonal. Again divide one part of Square into two parts. Thus we get two isosceles triangles.

Now place these two isosceles triangles on the opposite corners of the adjacent sides i.e. one on northern side and other on eastern side so that one rectangle is formed See 3.



CONSTRUCTION AND PROOF :

Draw a square ABCD. Join BD to divide it into two parts so that triangle DAB and triangle DCB are obtained. Again divide triangle DBC into two parts such that two isosceles triangles are obtained triangle EDC and triangle ECB. Place triangle ECB on triangle FAB (eastern side of square ABCD) such that C on A, B on B, E on F. Place triangle EDC on triangle GDA such that D on D, C on A, E on G. Thus GDBF is a required rectangle (see 4).

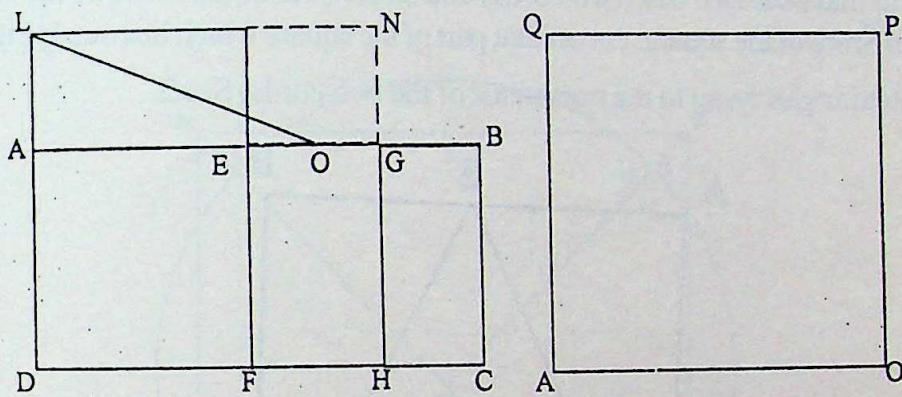
5. METHOD OF CONVERTING A RECTANGLE IN A SQUARE

दीर्घचतुरसं समचतुरसं चिकीर्षु रितिर्यद् मानी करणीं कृत्वा शेषं द्वेधा विभज्य पाश्वयोरुपदध्यात्
खंडमावापेन तत्संपूरयेन्तस्य निर्हार उक्तः ॥ BSS.1/54)

To transform ABCD a rectangle into square, draw a line EF bisecting the lines AB & DC. In this way AEFD and EBCF two squares are formed. Now in square EBCF, draw a line GH bisecting sides EB & FC. Therefore EBCF is divided into two rectangles

GBCH and EGFF. Place GBCH at AE and we have LMEA. Now complete square by placing small square MNGE above EGHF at corner. Thus LNHD is a big square. To convert it into a square whose area is equal to area of rectangle ABCD, we have to separate only the area of small square MNGE.

Take L as centre cut at AG equal to LN. Let it cut AG at O, complete a square with side AO. Hence AOPQ is a required square.



METHOD OF CONSTRUCTION AND PROOF:

To transform a rectangle into a square Baudhāyana suggests this following method.

ABCD is a rectangle whose side $BC = \frac{1}{2} AB$ (i.e. breadth is half of the length). Divide rectangle divided into two equal parts. Now other half is again divided into two parts. Its one half is GBCH which is eliminated from south and is placed on the east side of first half square AEFD. Now complete a square by introducing a small square at the corner so that LNHD is obtained. Now a/c to given condition

Area of rectangle ABCD = area of Sq. LNHD-area of Sq. MNGE.

Now with centre L and radius equal to LN Cut off AG at O. Join LO. The square on AO is equal to the difference between two squares.

$$AO^2 = OL^2 - AL^2$$

$$AL^2 + AO^2 = OL^2 \quad \text{and} \quad OL = LN$$

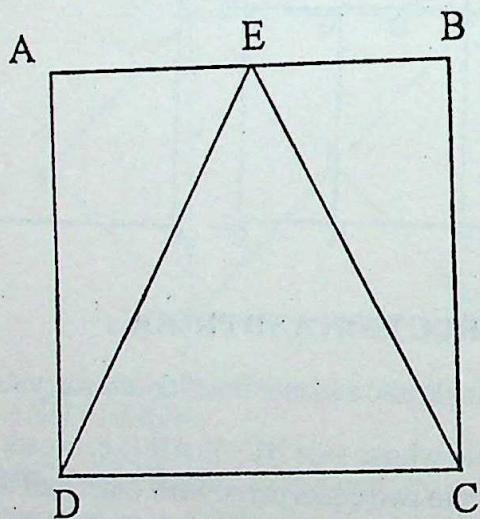
$$AO^2 = LN^2 - AL^2$$

Therefore, draw a square whose side is AO. Hence, new square whose area will be equal to rectangle ABCD is QPOA See 4.

6. TRANSFORMATION A SQUARE INTO A TRIANGLE

चतुरसं प्रउगं चिकीष्णावच्चिकीर्षेद् द्विस्तावर्तीं भूमि समचुतुरस्त्रां कृत्वा पूर्वस्याः करण्या
मध्ये शङ्कुं निहन्यात्तरिमन्पाशौ प्रतिमुच्य दक्षिणोत्तरयोः श्रोण्येर्निपातयेद्वहिः स्पन्द्यमपच्छिनद्यात् ॥
(BSS.1/56)

If you wish to turn a square into a triangle, make a square the area of which is twice as large as the area of the triangle you wish to make. Fix a pole in the middle of the eastern side, faster at that pole two ties (two cords) and stretch the cords towards the southern and northern Śroni of the square; cut off that part of the square which lies outside the cords (i.e., the two triangles lying to the north-east of the two cords) See 2.



CONSTRUCTION AND PROOF :

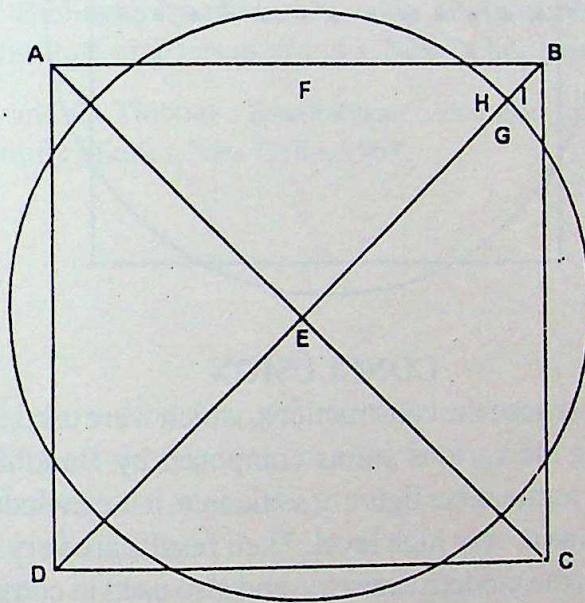
Draw a square ABCD. 'E' is the middle point of the side AB. Join ED and EC, cut off part ECD of the square which lies between the two triangles EDA and ECB.

Hence, ECD is a required triangle whose area will be half of the square ABCD
See 4.

7. TRANSFORMATION A SQUARE INTO A CIRCLE

चतुरसं मंडलं चिकीष्णायार्धं मध्यात्प्राचीमम्यापात यद्यदति—शिष्टते तरस्य सह तृतीयेन मंडलं
परिलिखेत् ॥ (BSS.1/58)

If you wish to turn a square into a circle, draw half of the cord stretched in the diagonal from the centre towards the *prāci*-line (i.e., stretch a cord from the centre of the square to one of the corners, for instance to the north-cast corner and move then the loose end of the cord towards south until the cord covers the *prāci* the line running from the centre of the eastern side of the square to the centre of the western side; a piece of the cord will then, of course, lie outside the square), describe the circle together with the third part of that piece of the cord which stands over (i.e., take for radius of the circle the whole piece of the cord which lies inside the square plus the third part of the piece which lies outside) See 2.



CONSTRUCTION AND PROOF :

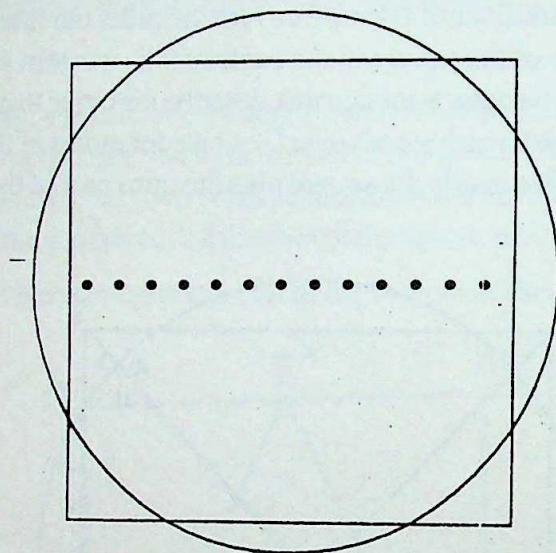
Transform ABCD square into a circle. Draw diagonal AC, BD from angle A and angle D. E is their meeting point, cut half of AB at F, cut off arc EG=FB from E. Divide GB into three parts GH, HI, IB. Now leave two parts and with centre EH draw a circle whose area will be equal to that square See 4.

8. TRANSFORMATION A CIRCLE INTO A SQUARE

मंडलं चतुरस्रं चिकीर्षन्विष्कम्भमष्टौ भागान्कृत्वा भागमेकोनत्रिप्रशाधा विभज्याष्टावि
शतिभागानुद्वरेद् भागस्य च षष्ठमष्टमभागोनम् ॥ (BSS.1/59)

If you wish to transform a circle into a square, divide its diameter into 8 parts and then divide one of these parts into 29 parts and subtract 28 of these and also the sixth part (of the proceeding sub-division) less the eighth part (of last).

The meaning is : $7/8 + 1/8.29 - 1/8.29.6 + 1/8.29.6.8$ of the diameter of a circle is the side of a square the area of which is equal to the area of the circle See 2.



CONCLUSION

Given above are some of the constructions, which were used in the preparation of altars. After discussing the various sutras composed by Baudhāyan regarding the conversion of one figure into another figure of same area, it is concluded that in Vedic age, geometric knowledge was at very high level. Their results are very much applicable as mathematical devices in the modern times too and also used in compounding of land in terms adding and subtracting land. These sūtras show that there was a concept of altering the different shapes of land into a desired one.

REFERENCES

1. Archaeological Survey of India, Annual Report, (ASI, AR), 1920-21, P.17.
2. Bodhāyanācārya and Bhagiratha Prasada Tripathi 'Vagisastri' (Ch.Ed.): Baudhāyana Śulba Sūtram; Research Institute, Sampurnanad Sanskrit Vishavidyalaya, Varanasi, 1979.
3. Datta, B.: The Science of Śulba, a study in early Hindu Geometry; Calcutta University, 1932.
4. B. Datta and A.N. Singh : History of Hindu Mathematics (part I and II); Asia Publishing House, Bombay, 1962.

5. R.C. Gupta : Vedic Mathematics from the Śulba Sūtras; J. of Mathematics Education, P.29, Vol. 9, No. 2, July 1989, 1-10.
6. Nidhi Handa : Kātyāyana Śulba Sūtra and Modern Mathematical Interpretations of its Sūtras; Ph. D. thesis G.K.V., 2002.
7. Lashan Lal Jha (Ed.) : Līlavatī of Bhāskarācārya II; The Chowkhamba Vidyā Bhawan, Chowk Varanasi, 1961.
8. S.D. Khadilkar (Ed.) : Kātyāyana Śulba Sūtra; Vaidika Samsodhana Mandoler, Poona, 1974.
9. S.N. Sen : A Bibliography of Sanskrit works on Astronomy and Mathematics part I"; National Institute of Sciences of India, New Delhi, 1966.
10. D.N. Yajvan and G. Thibout : Baudhāyana Śulba Sūtram; The Research Institute Ancient Scientific Studies, New Delhi, 1968.

SIMILARITY OF PLANE FIGURES AND GEOMETRIC & GROUP THEORETIC STUDY OF CYCLIC QUADRILATERALS OF FIRST KIND

VINOD MISHRA* AND S. L. SINGH**

(Received 12.06.03 and in revised form 27.10.2003)

ABSTRACT

Similarity of plane figures is a very old concept. The approach is to study the valuable contributions by the Arabs, Babylons/Summerians, Chinese and Greeks with reference to the Indians at large and importance of the concept as such. Finally, we study cyclic quadrilaterals of first kind geometrically. Its group theoretic study concludes with a non-abelian group of eight permutations in which four permutations are equal to another four.

Section A

INTRODUCTION

The concept of similar plane figures which is of ritual origin and whose reference goes back to at least the period of the *Śatapatha Brāhmaṇa* (ca. 2000-250 B.C.) originated due to the oftentimes necessity to construct, during the sacrificial rites, an altar similar to but differing in area from another by a specific amount. In it, for instance, regarding different layer construction of the *Caturasara-śyenacit*, a primitive fire altar, is said that it "should have an area of seven and a half square *puruṣa* at the time of first construction. At the second construction, its area shall have to be $8\frac{1}{2}$ square *puruṣas*, at the third $9\frac{1}{2}$ square *puruṣas*. In the same manner the area of the *Agni* should be increased by one square *puruṣa* at each successive construction up to $101\frac{1}{2}$ square *puruṣas*" (without depriving it (the *Agni*) its due proportions, i.e., shape) [1]. Construction of a *vedi* 14 or $14\frac{3}{7}$ times as large as the *Mahāvedi* and falcon-shaped *vedis* are the other examples to its credit (*op. cit.*, pp. 158-160). An elaborated collections and the other details as regards the construction of similar figures pertaining to various altars are available in the manuals, called the *Śulbasūtras* (ca. 800-200 B.C.). In these, for example, "the *Sautrāmaṇikī-vedi* is stated to be equal to one third the *Mahāvedi* and the *vedi* of *Asvamedha* double the latter. The *Lokṣahoma-vedi* and *Kotihoma-vedi* are respectively four and twenty-five times the *Pakayajniki-vedi*" [2]. A set of similar transformations for circles, rectangles, rhomboids, trapezia, triangles etc. may be found in

*Sant Longowal Institute of Engg. & Tech., Longowal.

**Gurukula Kangri Vishwavidyalaya, Haridwar. (E-mail : vedicmri@sancharnet.in)

the *Sulbasūtras* thus arising the propositions, a few popular among them are discussed here.

1. To construct an isosceles trapezium similar to a given one but of n times the size of it. The principle underlying is to replace the unit of measurement of the latter by \sqrt{n} times (*Ibid*, p. 154).
2. To construct a fire-altar similar to that of the shape of a falcon, but differing from its primitive area by m square unit (*Ibid*).
3. To construct a trapezium, provided one side (or altitude) of former is known (*Ibid*, p. 163).
4. To construct a rectangle similar to a given rectangle by successively varying the units (*Ibid*).

THEOREM OF SQUARE ON HALF CHORD

The *Āryabhaṭiya* (=AB) (c. 499 A. D.) of *Āryabhaṭa* regarding the above theorem reads as: In a circle (when a chord divides it into two arcs), the product of the arrows of the two arcs is certainly equal to the square of half the chord [3]".

Mathematically speaking, if in a circle (Figure 1) a chord CD is bisected at E by a diameter AB perpendicular to it, then

$$(1) \quad AE \cdot EB = CE^2.$$

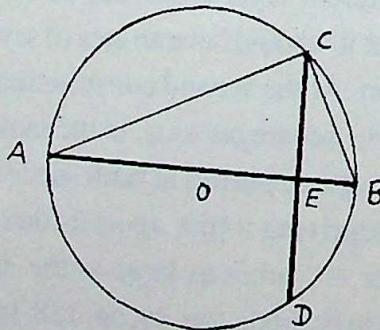


Figure 1

Its proof easily follows from the similar triangles CAE and CEB (see [4], p. 53). The flavour of this theorem is even embodied in the early Jaina works [5] as the *Jambūdvipasmāsa* and the *Tattvārthādhigama-sūtra* composed of Umasvati of first century A. D. atleast, a formal translation of latter which correlates the chord (c), the depth (h), and the diameter (d) of a circle runs as: "The square-root of four times the product of an arbitrary depth and the diameter diminished by that depth is the chord" (iii, 11). That is $c = \sqrt{4h(d-h)}$.

Trustworthy evidence of the late compositions [6] where the rule is said to be recurring in one form or the other include, for instance, *Bṛhatkṣetrasaṁśāsa* of Jinabhadra Gaṇi (ca. 609 A.D.), the *Brāhmaśphuṭasiddhānta* (=BrSS) (c. 628 A.D.) of Brahmagupta, the *Līlāvatī* (=LV) of Bhāskara II (b. 1114 A.D.), the *Naṭadapurāṇa* (=NP) (ca. twelfth century A.D.) and also *Ganita Kaumudi* of Nārāyaṇa Pandita (ca. 1356 A.D.).

Here are a few interesting examples based on the rule (1).

Problems from Bhāskara I's commentary the *Āryabhaṭīyabhāṣya* (=ABB) (c. 629 A.D.) on the diameter *AB*

I. Broken bamboo problem

We cite here one of the two similar examples proposed by Bhāskara I.

A bamboo of height 18 is felled by the wind. It falls at (a distance of) 6 from the root (thus) forming a triangle. Where is the break? [7]

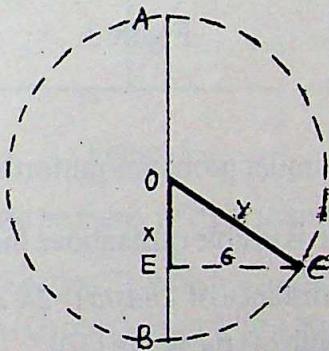


Figure 2

A few other references [8] of the similar problems are :

1. The *Ganitasāraṅgraha* (=GSS) (c. 850 A.D.) of *Mahāvīra*
2. The commentary (c. 860 A.D.) on the BrSS by *Prthudakasvami*
3. The *Bijaganita* and the *LV* of *Bijagaṇita* and the *LV* of *Bhāskara II*
4. The commentary on the *AB* by *Someśvara* (fl. 916-1200 A.D.)
5. The *Ganitakaumudi* (=GK) (c. 1356 A.D.) of *Nārāyaṇa Pandita*
6. The commentary on the *AB* (=ABRR) by *Raghunātha Rāja* (ca. 1597 A.D.)

Noticeable that the Chinese text *Chiu-chang suan-shu* (Nine Chapters on Mathematical Art, ca. 100 B.C.) too contains a problem similar to *Bhāskara* I in its ninth chapter entitled "kou ku" (Right Angled Triangles). Its solution is understood to be based on the Pythagorean property; Problem 13.

"A bamboo is 1 *chang* (= 10 *chhih*) tall. It is broken, and the top touches the ground 3 *chhih* from the root. What is the height of the break"? [9]

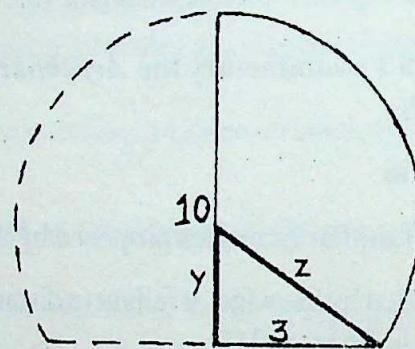


Figure 3

II. Lotus problem

Here is one of the two similar problems put forth by *Bhāskara* I.

"A full-blown lotus 8 *āngulas* is visible (just) above the water. When carried away by the wind, it submerges just at the distance of 1 *hasta* (=24 *āngulas*). Tell quickly (the height of) the lotus plant and (the depth of) the water [10]."

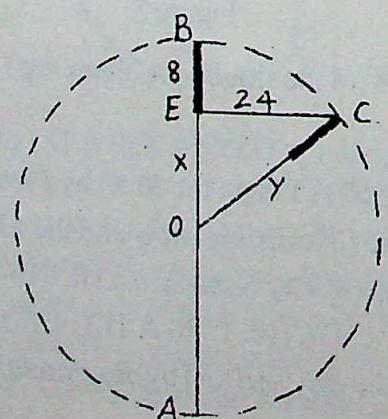


Figure 4

The similar problems [11] occur also in the *LV* and the *GK*.

Worth mentioning that one such problem similar to *Bhāskara* I we also find as the "problem of reed in a pond" in the ninth chapter of the *chiu-chang suan-shu*; Problem 6.

"There is a pond whose section is square of side 1 *chang* (= 10 *chihih*). A reed grows in its centre and extends 1 *chihih* above water. If the reed is fulled to the side (of the pond), it reaches the bank precisely. What are the depth of the water and the length of the reed [12]?"

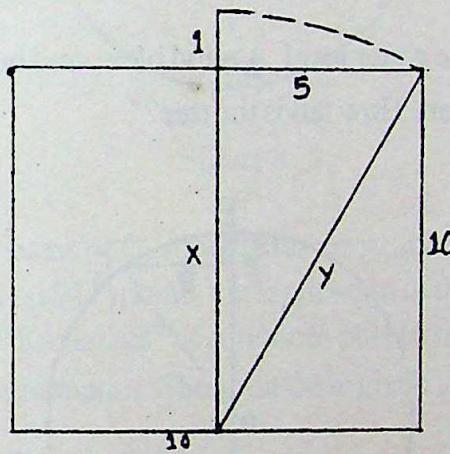


Figure 5

Certainly its solution is based on the property of right triangle.

In the Collection (Book III, Section 2), which is a commentary of Elements, Pappus (third century A.D.) of Alexandria (Greece) deals with a problem to show in a semi-circle the three means of arithmetic, geometric and harmonic. Following Heath [13], OC, BD & DF respectively stand for arithmetic mean, geometric mean and harmonic mean between two numbers AB and BC.

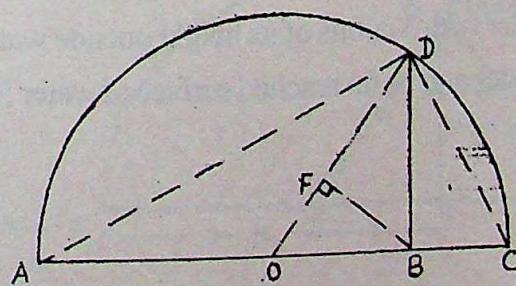


Figure 6

To mention here in this aspect that arithmetic and the theory of music led early Pythagorean School to develop the three means. The third mean now called harmonic due to Arahytas and Hippasus was then called 'subcontrary'; the other two were known as these are known presently [14].

Problem 42 [15] of the Persian arithmetic treatise *Miftāh al-muāmalāt* ("Key to Transactions") by Iranian mathematician and astronomer Muhammad bin Ayyūb Ṭabārī of 11-12th century A.D. [Ed. M. -A. Rīyahī, Tehran, 1970, in Persian].

"A tree 3 cubits was outside water level. A wind blew, and bent it so that its top reached the water level, 9 cubits apart. How tall is the tree?"

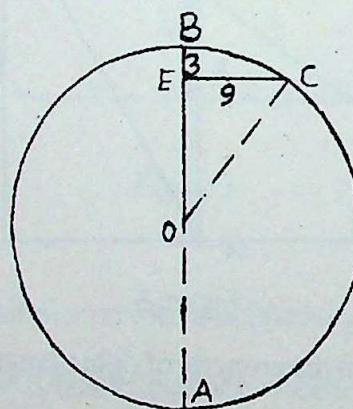


Figure 7

Ghiyath al-Din Jamshid al-kashi's *Miftah al-hisab* (first of the eight geometrical problems) [16] [Ed. A.S. Damardash and M.H. al-Hafni al Shaykh, Cairo, 1970].

"A spear is planted in water with 3 cubits of its height outside water. A wind inclined to it so that it submerged in water and its tip reached surface of water 5 cubits away. How long is the spear?"

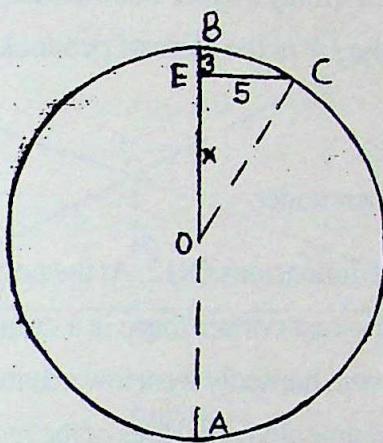


Figure 8

Bagheri [15] on the basis of Gardner [Mathematical Puzzles of Sam Lord, New York, 1959; Sam Lord 1841-1911] points the replication of the problem of *Bhāskara* II as 'Lily problem' in a novel "Kavanagh" of American poet Henry Wordsworth Longfellow (1807-1882), a lover of mathematics. The solution is given to be similar to Tabari.

II. Hawk and rat problem

One of the two problems from *Bhāskara* I's commentary is :

"A hawk is sitting on a pole whose height is 18 (cubits). A rat, who has gone out of his dwelling (at the foot of the pole) to a distance of 81 (cubits), while returning towards his dwelling, afraid of the hawk, is killed by the cruel (bird) on the way. Say how far has he to go to reach his hole (if he were to alive), and also the (diagonal) motion of the hawk (the speeds of the rat and the hawk being the same) [17]."

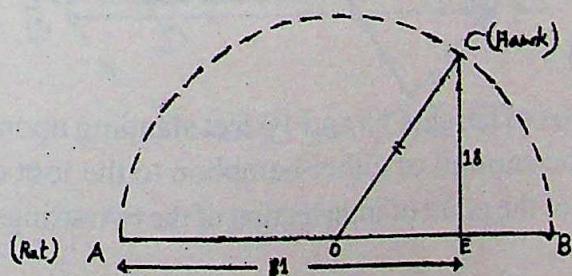


Figure 9

The same problem reappears in Yallay's super commentary [18] on the *AB*. The similar problem is stated to occur in the *LV* in the form of peacock and serpent pair [19].

IV. Crane and fish problem

The problem from the *ABB* is hereunder.

"There is a reservoir of water of dimensions 6×12 . At the north-east corner of the reservoir there is a fish; and at the north-west corner there is a crane (*bake*). For the fear of the crane the fish, crossing the reservoir, hurriedly went towards the south in an oblique direction but was killed by the crane who came along the sides of the reservoir. Give out the distances travelled by them (assuming that their speeds are the same) [20]."

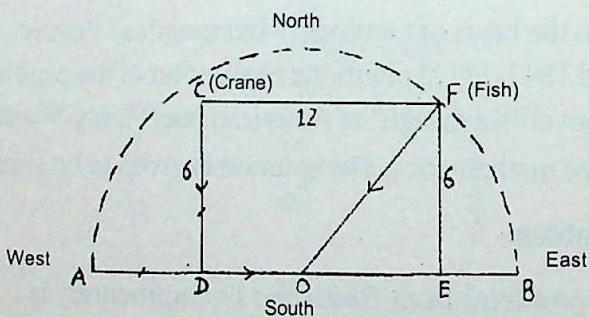


Figure 10

The same problem is claimed to be reproduced in the *GK* and the *ABRR* [21].

PROBLEM FROM THE *LILAVATI*

Bamboo problem [22].

"There are two bamboos of heights 15 and 10 feet standing upon the ground and tight strings are tied from the summit of either bamboo to the foot of the other. Find the perpendicular distance of the point of intersection of the two strings from the ground."

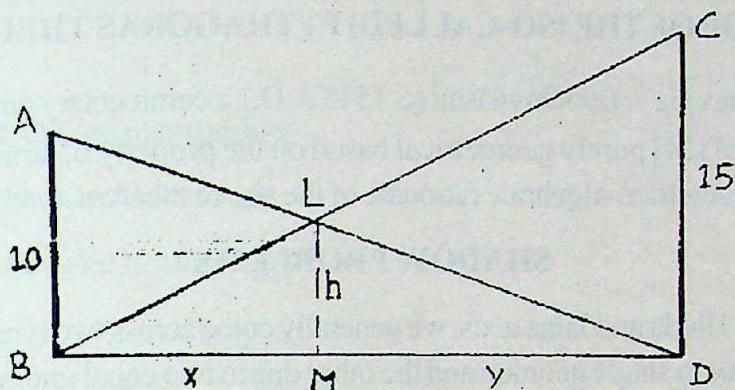


Figure 11

BRAHMAGUPTA'S THEOREM

Statement [23]

"The rectangle contained by the two sides of a triangle is equal to the rectangle contained by the circum-diameter and the altitude to the base." *BrSS XII.27.*

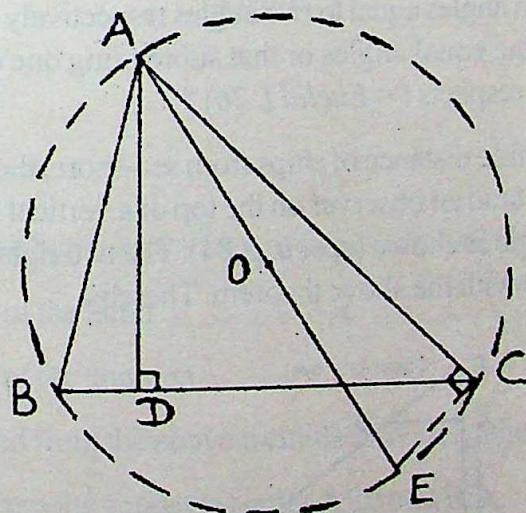


Figure 12

The result $AB \cdot AC = AD \cdot AE$ easily follows from the similar triangles ACE and ADB.

PROOF OF THE (SO-CALLED) PYTHAGORAS THEOREM

Ganesa Davjñā's *Buddhivilāśini* (c. 1545 A.D.), a commentary on the *LV*, attempts to furnish a proof [24] purely geometrical based on the property of similar triangles for Bhaskara II's geometricto-algebraic rationale of the above theorem.

SHADOW PROBLEMS

In ancient Hindu and Jaina texts, we generally come across two types of the shadow problems, one due to single gnomon and the other due to two equal gnomons (each based on the property of similar triangles). The latter appears in the *AB*, the *BrSS*, the *LV*, the *NP* and the *GK*, and also in the various commentaries on the *AB* by *Bhāskara I*, Suryadeva *Yajvān* (b. 1191 A.D.), *Paramesvara* (fl. 1380-1460 A.D.), Yallaya (ca. 1470 A.D.) and Raghunatha *Rāja*, whereas the former recurs in the various texts, viz. the *AB*, the *BrSS*, the *GSS*, *Sripati's Siddhāntasēkhara* (c. 1039 A.D.), the *LV*, the *NP* and the *GK*. For details, refer to Mishra and Singh [25].

Here is the Greek point of view :

Thales' (ca. 624-547 B.C.) geometrical theorem (without description) [26]:

"If two triangles have two angles equal to two angles respectively and one side to one side (namely that adjoining the equal angles or that subtending one of the equal angles, the triangles are equal in all respects (= *Euclid I. 26*)."

To measure inaccessible distance of ships from sea-shore, the one of the possibilities that Thales used was to stand an observer on the top of a vertical tower and constructing a small triangle in a triangle as shown (*op. cit. p. 84*). The two right triangles ABC & CDE are similar in accordance with the above theorem. Thereby

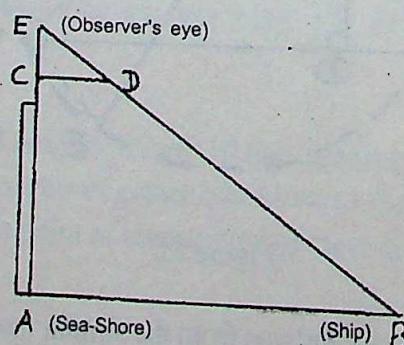


Figure 13

$$AB = \frac{AE \cdot CD}{CE}$$

for if AE, CD & CE are measurable.

Now we come back to Indian case.

Shadow problems due to single gnomon

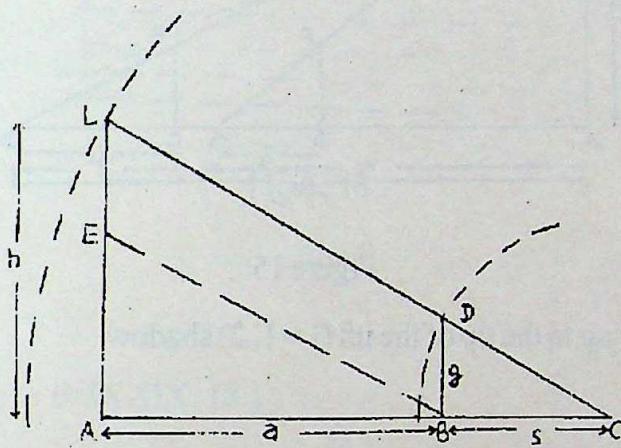


Figure 14

The Shadow-length

$$2) s = \frac{ag}{h - g}$$

wherein a = distance between the root of lamp-post and that of the gnomon

h = height of the lamp

g = height of the gnomon

[a and h can be evaluated from the above formula provided other parameters are known.]

The rationale follows from the similar triangles AEB and BDC.

Alternative explanation due to Nilkantha [27]: Source of light L is located on a circle concentric with the own circle (of gnomon and its shadow), C, the tip of the shadow, acts as centre and CD (D- the tip of the gnomon) as radius. Then LA, the vertical through L, is the sine-chord (*bhuja*) and AC (C- the foot of L) is the cosine-chord (*koti*).

Shadow problems due to two equal gnomons

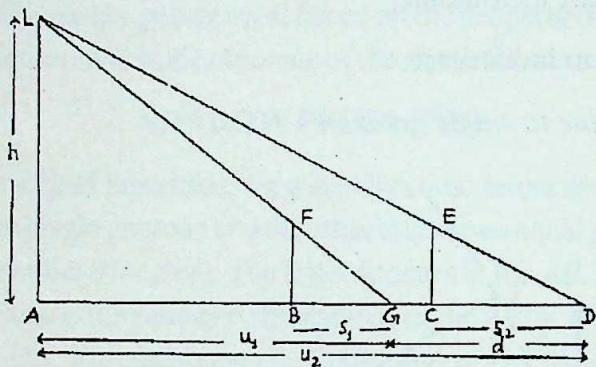


Figure 15

Upright corresponding to the tip of the i th ($i = 1, 2$) shadow

$$3) \quad u_i = ds_i / \Delta s.$$

Height of the lamp-post

$$4) \quad h = g(u_i/s_i)$$

$$\Rightarrow h = g(d/\Delta s); \quad \Delta s = s_2 - s_1, \quad \Delta u = u_2 - u_1 = d.$$

The results are easily derivable using the properties of similar triangles. However *Paramesvara*, a commentator of the *Āryabhaṭīya*, bases his rationale on the rule of three. Quite interesting problems and their exercises occur almost in all the above texts. However, we cite here only some enlightening (special) problems, other than the normal, from the *BrSS*, *LV*, *GK* and *GSS*.

I. Solutions to water-reflection problems [28]

(a) *To find height and distance of the object by observing its reflection from a distance*

"The distance between the house and the man is divided by the sum of the heights of the house and the man's eyes and multiplied by the height of the eyes. The tip of the image of the house will be seen when the reflecting water is at a distance equal to the above product" [BrSS XIX.17].

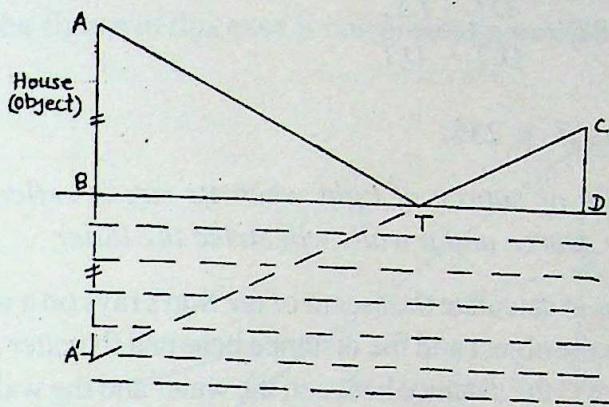


Figure 16

$$DT = \frac{BD \cdot CD}{AB + CD}$$

$$AB = \frac{BT \cdot CD}{DT} \quad [\text{due to BrSS XIX. 18.}]$$

For rationale, refer to [op. cit. p. 254].

(b) *To find height and distance of the object by observing reflection from two different distances*

"The distance between the first and second positions of the water divided by the difference between the distances of the man from the water, when multiplied by the height of the eyes, is the height, and the same, when multiplied by the distance between the water and the man, is the distance between the water and the house" [BrSS XIX. 19].

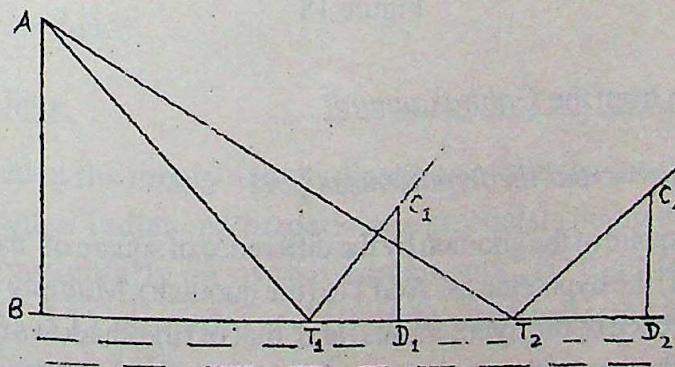


Figure 17

$$AB = \frac{C_1 D_1 \cdot T_1 T_2}{D_2 T_2 - D_1 T_1}, \quad BT_1 = \frac{D T_1 - T_1 T_2}{D_2 T_2 - D_1 T_1}.$$

Consult for proof, *op. cit.*, p. 255.

(C) *To find the height or source of light when its ray is reflected from a water-surface (between the source and a wall) will strike the latter*

"He... who knows how to calculate the ascent of the Sun's rays on a wall from the known ratio of the shadow to the object and the distance between the water and the wall" [BrSS XIX. 8]. (In such a case) "the distance between the water and the wall divided by ratio of the shadow to the object is the height of ascent" [BrSS XIX.20].

$$AB = \frac{BT}{DT/CD}. \quad \text{See } [op. cit., \text{ p. 256}] \text{ for its rationale.}$$

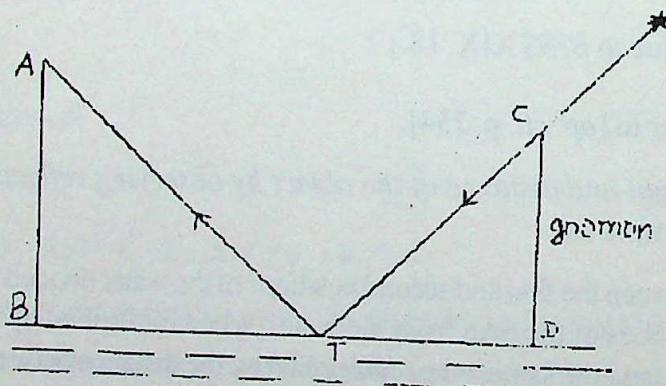


Figure 18

II. Shadow problem from the Ganitakaumudi

Lengths of the shadows and the hypotenuses [29]

"Divide 4 times the square of the gnomon by the difference of square of 'the difference of the shadow's and (that of) the hypotenuses'. Add 1 to (the quotient). Multiply the square-root of the sum by the difference of the hypotenuses (and, that of) the shadows (separately). (The products) happen to the sum of the shadows (and that of) the hypotenuses, respectively. The shadow and the hypotenuses should be obtained from them by *samkrarama*. The heights of the lights should be obtained by the method stated earlier" [GK 17-18].

[This problem seems to be an extension of the LV 232 which restricts to find shadow-lengths only. And the figure in this case is confined to gnomon, shadow-lengths and hypotenuses.]

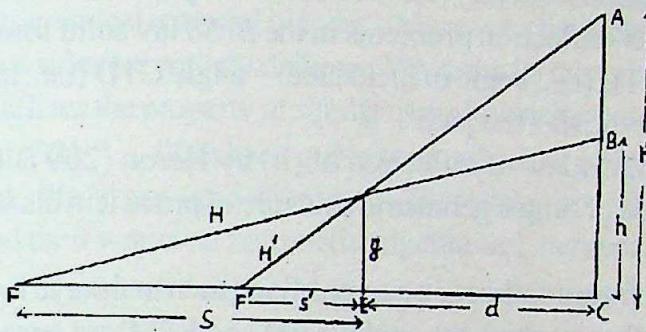


Figure 19

Let $s - s' = a$, $H - H' = b$, then according to the rule

$$s + s' = \sqrt[b]{\frac{4g}{a^2 - b^2}} + 1 \text{ and } H + H' = \sqrt[a]{\frac{4g}{a^2 - b^2}} + 1.$$

Then by *samkramana*

$$s \text{ or } s' = \frac{1}{2} \left[\pm a + b \sqrt{\frac{4g^2}{a^2 - b^2} + 1} \right], \text{ and}$$

$$H \text{ or } H' = \frac{1}{2} \left[\pm b + a \sqrt{\frac{4g^2}{a^2 - b^2} + 1} \right].$$

$$\therefore h = dg/s \text{ and } h' = dg/s'.$$

III. The nature of light

The light ray used in the theory and applications of the shadow problems certainly reflect the wisdom of Indian mathematicians/physicists being convergent with the nature of light particularly on its two aspects, viz., (i) light travels in a straight line (fifth century A.D. onwards), and (ii) light when reflected follows the shortest path (seventh century A.D. onwards).

Explanation

I am rather inconclusive whether the *Vedic* people were aware of the fact (i), but certainly after the dark age, in the first written treatise on mathematics of elder *Aryabhata*, the shadow problems exhibit it silently (for instance, the ray LDC is a straight line, Fig. 12). Again the three water-reflection problems in the *BrSS* lay solid foundation to (ii), for instance, by angle ATB (i.e., angle of incidence) = angle CTD (i.e., angle of reflection) (Fig. 16) is meant for the shortest path.

Below is a proof³⁰ of the law of reflection of (ii) by Heron (200 B.C.) of Alexandria (Greek) who made use of simple geometric structure of prove it in his work on *Catoptrics* (i.e., reflection):

"If a light is to travel from a source to a mirror M' and then to the eye E of an observer, the shortest path SPE is that in which the angles SPM and EPM' are equal" [cf. 30].

Let the other path $SP'E$ be feasibly short. Let $SQS' \perp MM'$ such that

$SQ = S'Q$. Join S' to P and P' . The path-lengths

$SPE = S'PE$ & $SP'E = S'P'E$.

Of the two, $S'PE$ is a straight line (because angle $M'PE$ = angle $M'PS$) and is, therefore, shorter than the bent line $S'P'E$. Hence path $SPE <$ path $SP'E$.

The knowledge of (ii) is even credited to Euclid and Aristotle and sometime also to Plato (*op. cit.*, p. 194).

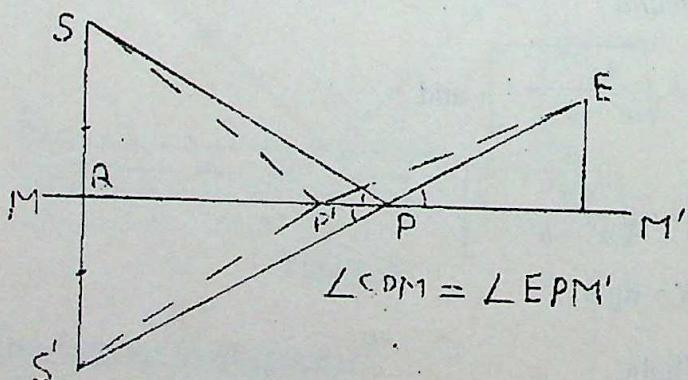


Figure 20

CONCLUDING REMARKS

Various studies reveal that Buddhism was the main vehicle for transmission of Indian culture to China. During the intercourse, a large number of Indian works were translated into Chinese language between first and early eighth century A.D.

The parallel problems on broken bamboo, lotus among others in India throw some light on the possible Chinese transmission to India. However, according to Gupta [31], ".... while things were documented in Chinese sources, there is no familiar positive literary or other documentary evidence known from Indian sources which specify clearly the arrival of any Chinese mathematical material in India". Moreover, the techniques employed for solution from the two sides are entirely different. While the former is based on Pythagorean property, the later utilizes the property of similar triangles arising out of a chord in a circle. According to Maiti [32], "... *Bhāskara* quotes one *gāthā* in *prākṛta*, which states the rule for the chord, the arrow and diameter. Therefore, his reference to previous mathematicians and their works on arithmetic, algebra and mensuration during or before *Āryabhaṭa* leads us to assume that *vāniśabhaṅga* problem may have been devised by either *Maskari* or *Pūrana* or *Mudgala* or *Putana*. And if *Maskari*, the mathematician be the Makkhali of the Buddhists or Mamkhali of the Jainas or *Maskari* of Panini and *Pūrana* *Kassapa* or *Pūrana* of the *Mahābhārata*, then *śyenamūsika* problems along with *vāniśabhaṅga* may have their origin in the Buddha's time, at least. This may not be an impossible probability because certain numeral terms are found to occur in *Maskari*'s creed and these may have been made use of, in all probability, in the pursuit of his system of principle. Further, measurements with bamboo were extensively known during *Brāhmaṇa* and *Śulba* periods. *Āpastamba* had made use of a bamboo-rod to construct a square having a given side. In the time of *Arthaśāstra* (400 B. C.) the length measure *danda* is found to have been in common use. An interesting story of measuring a Buddha statue with a bamboo pole had been referred to by the author of the *She-Kia-Fang-Che*: "In former times there was a man who with a sixteen feet pole tried to measure Buddha". These facts, although meagre, also allude to the probability of setting up the problem by Indian Mathematician themselves. The Jaina *Syādvāda* (Seven Fold Prediction) in which lies the seed of the theory of the probability also suggests India's mathematical attainment to a comparatively high degree in 400 B.C. ".

Noticiable that, as pointed out by Van der Waerden³³ (see Bag [11]), Babylonian text BM 34568 (1900-1600 B.C.) - according to his classification - gives 13 types of problems involving $x^2 + y^2 = z^2$, and that the bamboo problem comes under type VI : Given $x = a$, and $y + z = b$, to find x and z . More interestingly the four problems of the above Babylonian text are found to be identical with the Chinese type 1, 2, 3, and 6 [26]. After a detailed comparative study Van der Waerden "conjectures a common origin of mathematics" which he called pre-Babylonian mathematics (cf. Bag, ref. 33).

The remaining problems other than the above two enjoy the full credit of being Indian

contribution and free from any foreign influence.

The various geometrical problems concerning light confirm the Indian perception of the two basic laws of it certainly after the work of Brahmagupta. In this context the remark by T. A. Sarasvati ([28], p. 256)'... the possibility of Brahmagupta having investigated the laws of reflection' is worth noting.

SUPPLIMENT

In the *Śulbasūtras*, many foundational aspects of the knowledge of similar figures are practically observed due to altar construction. The one -" the corresponding sides and lines of similar figures are proportionate " - for example, has an interesting application in the construction of the bricks to be used in the wing of the *vakrapasa-śyenacit* (a falcon shaped altar with curved wings.) *Āpastamba* says, "The traverse side is $1/7^{\text{th}}$ of the side of the side of the wing and the lateral side is $1/4$ of a *puruṣa*. Its frame should be expanded diagonal-wise. The planks should be inclined by $1/7^{\text{th}}$ of the *pakṣanamāṇī* -the shape or gradient of the wing" [Sarasvati, [27], pp. 48-49].

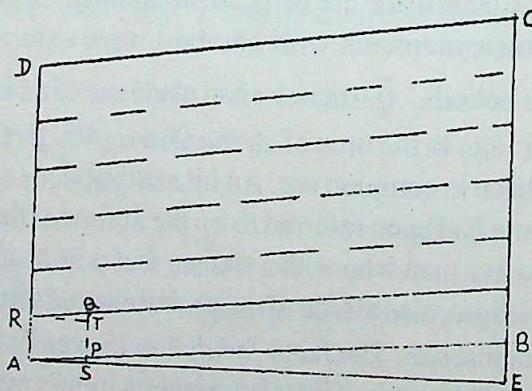


Figure 21

Here *pakşanamani* means for the vertical height of the tip of the inclined wing from the horizontal. ABCD is half of the wing as shown. CD is produced to meet the horizontal through A in E. BE is the *pakşanamani*, here. If APQR is a brick, QP is extended to meet the horizontal through A is S. The triangle APS is similar to the triangle ABE. Therefore

$$\frac{PS}{BE} = \frac{AP}{AB} = \frac{1}{7} \text{ or } PS = \frac{1}{7} BE.$$

Application of the rule (theorem of square on half chord) to find the radius of curvature (*vakratārdhyavyāsa*) of spherical surface

Corresponding to 'a', half of the chord (*jivādharm*), and 'h', arrow (*sāra*), the radius of the circle (*vṛta*) will be governed by the formula

$$R = \frac{a^2}{2h} + \frac{h}{2}$$

To find the radius of curvature of a spherical surface, we need to compute, 'a' and 'h' with perfection, for which in modern days, an instrument called *spherometer* is used. A spherometer basically consists of three equal equilateral legs and a central leg capable of moving vertically with the help of metallic circular plate. Very firstly the tip of the central leg is brought in contact with the plane glass plate and distance between the tips of central and outer legs, 'a' is measured. It is then placed gently on the spherical surface rotating slowly the screw so as to move central leg upward till three outer legs are just in touch with the surface, the net upward movement 'h' is measured from the graduated scale and the marked metallic disc. The final setting of the spherometer and the spherical surface is as shown in the Figure. [N. G. Dongre, *Physics in Ancient India*, New Age International Publishers, New Delhi, 1994, 48-49].

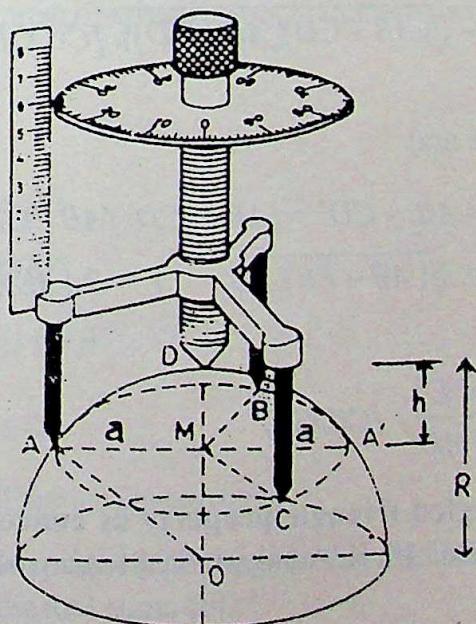


Figure 22

Algebraic proof of eq. (1) (as in the *LV*) [K.S. Patwardhan, S. A. Naimpally and S. L. Singh, *Lilāvatī of Bhāskarācārya*, Motilal Banarsi Dass Publishers, Delhi, 2001, pp. 146- 147]

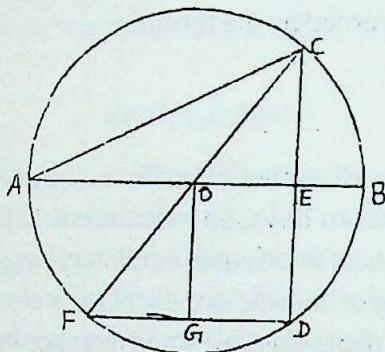


Figure 23

$$CO = OB.$$

$$\Rightarrow CO - OE = OB - OE = EB.$$

$$\therefore AE = AO + OE = CO + OE = \frac{CO^2 - OE^2}{CO - OE} = \frac{CE^2}{EB}.$$

$$EB = OB - OE = \frac{1}{2}[AB - \sqrt{(AB + CD)(AB - CD)}]. \quad [CCXIII, 1st \text{ half}]$$

$$[\text{In it, } OE = GD = \frac{1}{2} FD, \text{ and}]$$

$$FD^2 = CF^2 - CD^2 = AB^2 - CD^2 = (AB + CD)(AB - CD).]$$

$$CD = 2CE = 2AE. \quad EB = \sqrt[2]{(AB - EB) \cdot EB} \quad [= 2\sqrt{AE \cdot (AB - AE)}]$$

[CCXIII, 2nd half]

$$AB = AE + EB = AE + \frac{CE^2}{AE}. \quad [CCXIV]$$

Application of right-angled triangle property as conceived in old Babylonian/Summerian of about 1900 B.C. [S. K. Adhikari, Babylonian mathematics, *Indian J. Hist.* Sc. 1998, 33, pp. 17-18].

From a Babylonian cuneiform text (the text name is not mentioned by Adhikari):

To find the length of the chord of a circle when diameter and depth of the sector is respectively ($d = 20$) and ($h = 2$).

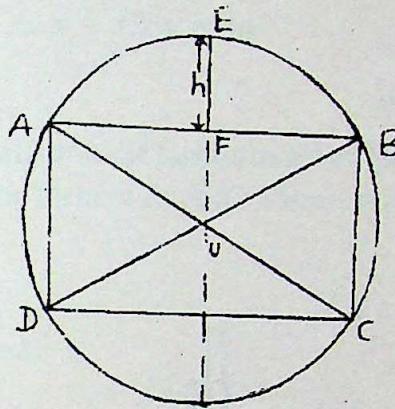


Figure 24

$AB^2 = AC^2 - BC^2$ [$= d^2 - [d - 2h]^2 = 4(d - h)h$] is the same as eq. (1).

$AC = d = 20$, $BC = d - 2h = 20 - 4 = 16$.

$$\therefore AB = \sqrt{AC^2 - BC^2} = 12.$$

The method of solution as prescribed can be put in modern form as :

$$2 \times 2 = 4$$

$$20 - 4 = 16$$

$$20 \times 20 = 6(40) [= 400]$$

$$16 \times 16 = 4(16) [= 256]$$

$$6(40) - 4(16) = 2(24) [= 48]$$

$$\sqrt{2(24)} = 12 \text{ is the chord.}$$

From a Summerian tablet (Thureau Dangin text)

"A straight rod 30 in. stands upright then it slips so that the upper end slides 6 in. down a vertical wall. How far has the foot come out?"

In Babylonian way the solution can be put as:

Patu 30

$$30 - 6 = 24$$

$$24 \times 24 = 9(36) [= 576]$$

$$15 - 9(36) = 5(4) [=324]$$

$$\sqrt{5(24)} = 18$$

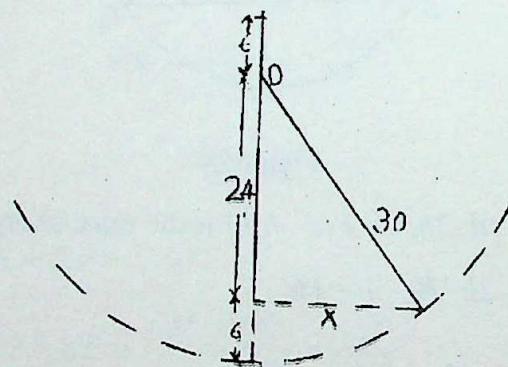


Figure 25

In the Figure,

$$x = \sqrt{(d-h)h} = \sqrt{(60-6)6} = \sqrt{324} \equiv 18.$$

Note. A number within [] means its modern equivalent.

BRIEF SOLUTIONS TO EXAMPLES ON THE OPEN MARKET

Broken bamboo problem

$$\text{Eq. (1)} \Rightarrow EB = CE^2 / AE = 36/18 = 2$$

$$x + y = 18, y - x = 2 \Rightarrow x = 8, y = 10$$

Parallel Chines problem

$$y^2 = z^2 - 32 = (y - 10)^2 - 9 \Rightarrow y = 91/20 = 4\frac{11}{20}.$$

Lotus problem

$$\text{Eq. (1)} \Rightarrow AE = CE^2 / EB = 576/8 = 72.$$

$$x + y = 72, y - x = 8 \Rightarrow x = 32, y = 40.$$

Tabari's problem

As reported, his solution in algorithmic form is based on the property of intersecting chords in a circle and is given in the form of Euclid's *Elements Book III (Theorem 34)*, without any explanation.

$$9 \times 9 = 81$$

$$[CE \times CE = CE^2]$$

$$81 : 3 = 27$$

$$[CE^2 / EB = AE]$$

$$27 + 3 = 30$$

$$[AE + EB = AB]$$

$$30 : 2 = 15$$

$$[AB/2 = OB]$$

Al-Kāshī's problem

According to *al-Kāshī* (following *Tabari*) [cf. 15]

"Square of CE divided by EB gives EA : 5 x 5 : 3 = 25 / 3. Adding EB gives BA = 34/3. This is the diameter, and by halving it the radius is found, which equals the length of the spear : 34/3 : 2 = 17/3 = 5 + 2/3 cubits."

Al-Kāshī does not hesitate to present algebraic methodology for its solution which is as follows [cf. *op. cit.*]:

$$EO = x, x^2 + 25 = CO^2 = OB^2 = (x + 3)^2$$

$$x^2 + 25 = x^2 + 6x + 9, 6x = 16, x = 8/3$$

$$8/3 + 3 = 17/3 = 5 + 2/3.$$

Parallel Chinese problem

$$x = (z^2 - e^2) / 2e = (25 - 1) / 2 = 12.$$

Hawk and rat problem

$$\text{Eq. (1)} \Rightarrow EB = CE^2 / AE = 324 / 81 = 4.$$

$$CO = 42 \frac{1}{2}, OE = 38 \frac{1}{2}.$$

Crane and fish problem

$$EB = FE^2 / AE = FE^2 / (CD + DE) = 36 / 18 = 2.$$

$$OF = AO = AB/2 = (AE + EB)/2 = (18 + 2)/2 = 10.$$

Bamboo problem

$h = 6$ feet follows from the similar triangles BCD & BLM and ABD & LMD.

$$h/15 = x/(x + y) \text{ and } h/10 = y/(x + y) \Rightarrow y = \frac{3}{2}x.$$

SECTION B

INTRODUCTION

This part is written with the intention of John Stillwell that "... ancient mathematics is not dead or obsolete- it stays alive because its problems continues to stimulate the creation of new concepts and techniques. The most fertile problems in mathematics are over 2000 years old and still not yielded up all their secrets" (p.1, reference [4]). Indian mathematics is proud of his genius Brahmagupta (b. 598 A.D.) for beautiful construction of a cyclic quadrilateral with perpendicular diagonals from two pythagorean triplets. For details of further study in the field, refer to Mishra and Singh [3]. The intent of this part is to reveal the further facts utilizing the concepts of geometry and group theory when two adjacent sides of the prescribed quadrilateral corresponding to a vertex of circum-diameter are equal; the two other sides corresponding to the second vertex of the same diameter has to be different.

GEOMETRIC STUDY

Let (a_0, b_0, c_0) and (a, b, c) be two pythagorean triplets with $a_0 > b_0, a > b$. These triplets readily supply the sides of cyclic quadrilateral A'BC'D with perpendicular diagonals, according to Brahmagupta, as $C'D = a = a_0c$, $BC' = b = ac_0$, $A'B = \phi = b_0c$ & $DA' = d = bc_0$, taken in order. The diagonals are :

$$d_1 = AC = AL + LC = b_0b + a_0a.$$

$$d_2 = BD = BL + LE = ab_0 + a_0b.$$

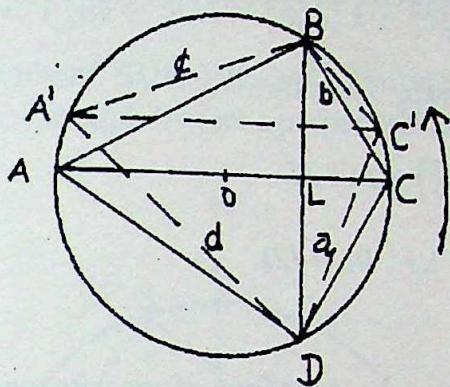


Figure 1a

This structure leads to the following theorem :

Theorem. In a cyclic quadrilateral ABCD with sides as above, the following statements are equivalent for $\rightarrow A' = A$, $\rightarrow C' = C$.

- Angle at B (and angle at D) = a right angle (each).
- $AB^2 + BC^2 = AC^2$ (and $AD^2 + DC^2 = AC^2$).
- $BL = LD$, i.e., $a/a_0 = b/b_0$.
- $BL^2 = AL \cdot LC$.

Proof. For (a) \Rightarrow (b) & (b) \Rightarrow (a), the proof is well-known and need no reproduction.

For (b) \Rightarrow (c), $AB^2 + BC^2 = AC^2$, i.e,

$$b_0(a^2 + b^2) + a^2(a_0^2 + b_0^2) = (bb_0 + a_0a)^2.$$

$$\Rightarrow ab_0 = a_0b, \text{ i.e., } BL = LD.$$

For (c) \Rightarrow (b), let $BL = LD$, i.e, $ab_0 = a_0b$ or $a/a_0 = b/b_0 = k$, say;

$\Rightarrow a = ka_0, b = kb_0$ & so $c = kc_0$; $\phi = d = kb_0c_0$ & $b = a = ka_0c_0$. Notice that $\phi < b$ since $b_0 < a_0$.

Now $AB^2 + BC^2 = k^2 b_0^2 c_0^2 + k^2 a_0^2 c_0^2 = (kb_0^2 + ka_0^2)^2 = AC^2$.

For (c) \Rightarrow (d), $a = ka_0$, $b = kb_0$. Therefore

$$BL^2 = (ka_0 b_0)^2 = (ka_0^2)(kb_0^2) = LC \cdot AL.$$

This implies that $ab_0 = a_0 b$, i.e., $BL = LD$.

This completes the proof.

Remark. The above figure may be in the form of opposite sense.

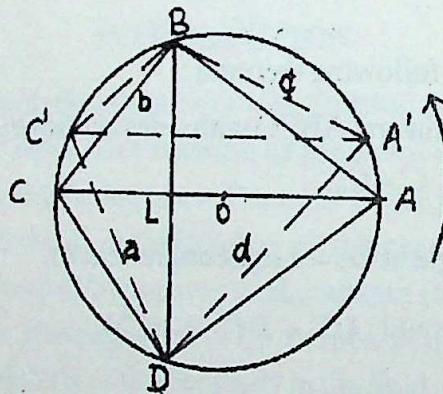


Figure 1b

Conclusion. In nut shell, we see that if (a_0, b_0, c_0) is a primitive triangle with $a_0 > b_0$, and k any scalar, then it is always possible to construct a cyclic quadrilateral with two adjacent sides $ka_0 c_0$ each through a vertex of the diagonal of the circle (i.e., diagonal of the quadrilateral) and other two sides $kb_0 c_0$ each through other vertex of the same diagonal. The circum-diameter is given by kc_0^2 , and the perpendicular diagonal by $2ka_0 b_0$.

Geometric construction otherwise and deduction therefrom

Figure 1a or 1b with four points ABCD possesses reflexive symmetry about AC. That is, any study of half-plane (semi-circle) will automatically reveal the secrets of the quadrilateral under consideration. The study of semi-circle by drawing a perpendicular from a point on half circle onto the diameter has several references in theory and applications in ancient mathematical world of Arab, Babylonia/Summeria,

China, India and Greek. The study had been done so far using Pythagoras theorem and similarity concept.

Let AL be a line segment. Extend AL to C . Letting O the midpoint of AC as centre draw a semi-circle. Through L erect a perpendicular BL on AC .

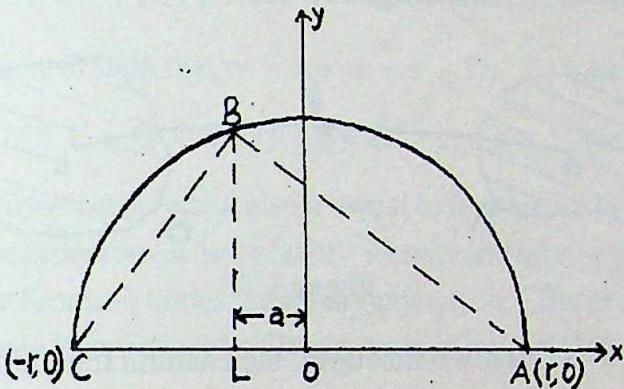


Figure 2

Equation of semi-circle through ABC

$$(1) \quad x^2 + y^2 = r^2, \quad y > 0.$$

Equation of line BL parallel to $x = 0$

$$(2) \quad x = -a.$$

Using (2) in (1),

$$(3) \quad BL^2 = r^2 - a^2 = (r+a)(r-a) = AL \cdot LC.$$

In particular if $LC = 1$, then $BL = \sqrt{AL}$. This implies that if $\alpha (= AL)$ is constructible then so is $\sqrt{\alpha} (= BL)$. Notice that "a real number α is said to be *constructible* if one can construct a line segment of length α in a finite number of steps from this segment of unit length by using a straightedge and a compass" [1, p. 354].

GROUP THEORETIC STUDY

In study of permutation groups, the defining terms of function, composite function, permutation and group are of utmost importance. For greater insight, these are explained briefly as hereunder. For the sake of completeness, we recall certain elementary definitions.

Assume A and B are two non-empty sets.

A *function* or *mapping* ϕ from A into B (symbolically $\phi : A \rightarrow B$) is a rule that assigns to each element a of A a unique element b of B. We say that ϕ maps a into b and is written as $\phi(a) = b$. Equivalently, b is the image of a under ϕ .

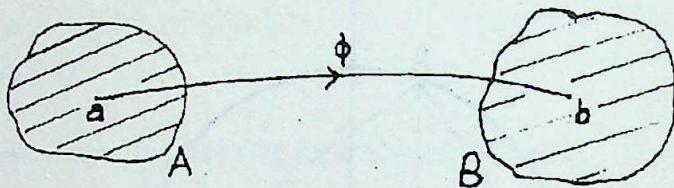


Figure 3

If $\phi : A \rightarrow B$ and $\psi : B \rightarrow C$ are two functions, then natural function mapping A into C consists of ϕ followed by ψ . For $\phi(a) = b$, $\psi(b) = c$; $(\psi \phi)(a) = \psi(\phi(a)) = \psi(b) = c$. $\psi \phi$ is a *composite function*.

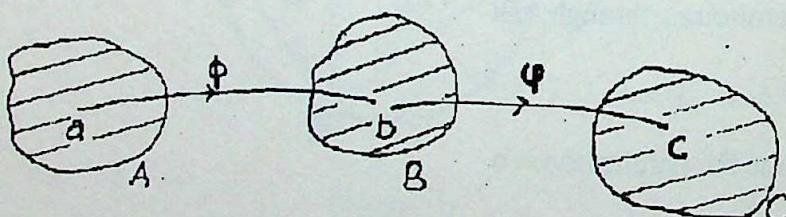


Figure 4

We call a function from A into B *one to one* (or *injective*) if each element of B has atmost one element of A mapped into it, and *onto* (or *surjective*) if each element of B has atleast one element of A mapped into it.

A *permutation* of a set A into itself is a *bijective map*, i.e. one to one function from A onto A. In this case, each element in the image appears once and only once. For instance, $\{2, 4, 5, 1, 3\}$ is one of the permutations of the set $\{1, 2, 3, 4, 5\}$ out of $5!$, but not $\{3, 2, 5, 4, 3\}$.

A *binary* operation on a non-empty set S (or a *composition* in S) is basically a function $\phi : S \times S \rightarrow S$. The image $\phi(a, b) \in S$ for $a, b \in S$ is usually denoted convincingly by ab or $a.b$ (multiplicative notation), $a + b$ (additive notation), a^*b etc. A binary operation on S is

equivalent to the statement: If $a, b \in S$ then $ab \in S$ (closure axiom).

A group G is an ordered pair (G, \cdot) consisting of a non-empty set G together with a binary operation on G satisfying the properties:

- i) $(ab)c = a(bc) \quad \forall a, b, c \in G;$ (associative law)
- ii) \exists an element $e \in G$ such that $ea = ae = a \quad \forall a \in G;$ (existence of identity)
- iii) for each $a \in G, \exists a' \in G$ such that $a'a = aa' = e.$ (existence of inverses)

Notice that the *identity element* e and also a' equal to *inverse* of a (denoted by a^{-1}) are unique. Eventually, the properties of associativity, identity and inverse altogether determine a group of one to one function under the binary operation. [The abstract definition of group is due to Leopold Kronecker (1870) and Arthur Claley (1878)].

A group G is said to be *commutative* iff its binary operation is commutative, i.e., $ab = ba$ for all $a, b \in G$. This group is sometimes called *abelian* group (named in honour of N. H. Abel (1802-1829) of Norway who did remarkable work in this field).

Theorem. Let A be a non-empty set and S_A the collection of all permutations of A . Then S_A is a group under permutation multiplication.

Mirror. In Figure 5, line m acts as a mirror (or a line of reflection), i.e., every point A of the plane on m acts as its own image, i.e., mapped to itself ($\sigma(A) = A$) under transformation σ . Such a point is the fixed point of σ . And the image of a point P off m to P' , $\sigma(P)$ on the opposite side of m . Eventually, m serves as a right bisector of the line PP' . Also distance and angle measure are preserved (i.e., $[R, S] = [R', S']$ and $\angle RST = \angle R'S'T'$ in reversing sense). In other words, every line m determines an injective map of the plane.

Mirror axiom. For every line m there exists a collineation that preserves distance and angle measure, fixes m pointwise, and interchanges the half planes of m .



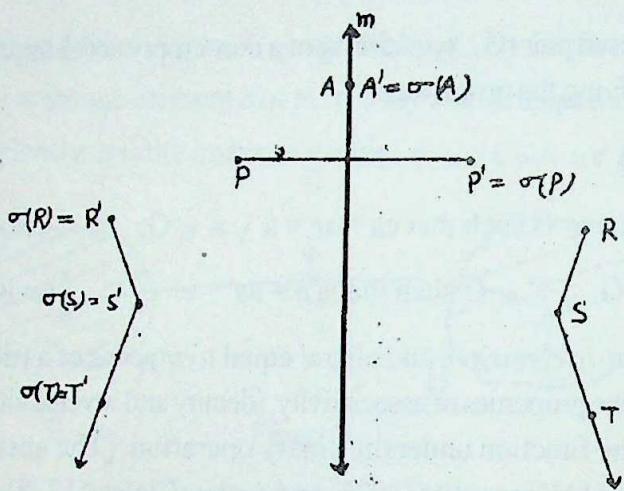


Figure 5

Assume set $A = \{\alpha, \alpha, \beta, \beta\}$, taken in order of 1 to 4 to avoid any mistake in further operations, denotes the side-lengths of a cyclic quadrilateral, where $\alpha = k\beta_0\gamma_0$ & $\beta = k\alpha_0\gamma_0$; $(\alpha_0, \beta_0, \gamma_0)$ is a Pythagorean triplet ($\alpha_0 < \beta_0$) and k a scalar. For a systematic study of group of permutations of A , we need to symbolize ρ_i for permutations corresponding to sides of Figure 6 and μ_j for permutations corresponding to the mirror image of this figure. The eight elements of S_8 , the collection of all permutations of A , are:

$$I = \rho_0(1234) = \begin{bmatrix} a & a & b & b \\ a & a & b & b \end{bmatrix} = \mu_2(2143)$$

$$\rho_1(2341) = \begin{bmatrix} a & a & b & b \\ a & b & b & a \end{bmatrix} = \mu_1(1432)$$

$$\rho_2(3412) = \begin{bmatrix} a & a & b & b \\ b & b & a & a \end{bmatrix} = \mu_4(4321)$$

$$\rho_3(4123) = \begin{bmatrix} a & a & b & b \\ b & a & a & b \end{bmatrix} = \mu_3(3214)$$

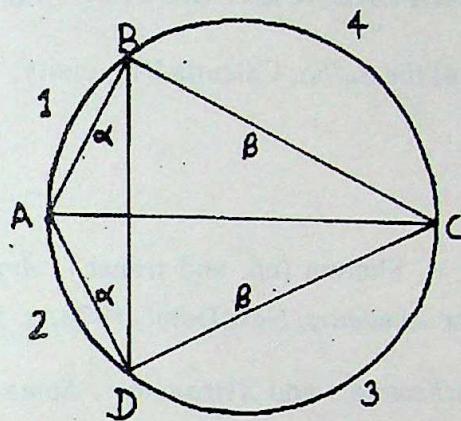


Figure 6

There is one to one correspondence among ρ_i s and μ_j s in a manner

$$\rho_1 \leftrightarrow \mu_1; \rho_2 \leftrightarrow \mu_4; \rho_3 \leftrightarrow \mu_3; \quad I = \rho_4 \leftrightarrow \mu_2$$

and is due to the fact of symmetry about circum-diameter AC. The choice of i and j in ρ and μ are as per convenience.

The credit for definition of group of permutations and 2- row notation goes to the French mathematician Augustin-Louis Cauchy (1789-1857).

Notice that, for instance,

$$\rho_1 \mu_3 = \mu_2 \mu_3 \rho_1 = \mu_4, \text{ i.e., } \rho_1 \mu_3 \neq \mu_3 \rho_1$$

$$[b] \rho_p \rho_q (A_i) = \rho_p (\rho_q (A_i)) = \rho_p (A_i) = A_k, 1 \leq i, j \leq 4].$$

Hence S_8 (octic group) is non-abelian.

REFERENCES (SECTION - A)

1. Datta, B., *The Science of the Sulba*, Calcutta University, 1932 (reprinted: 1991), p. 152.
2. *Ibid.*
3. Shukla, K. S. and K. V. Sharma (ed. and transl.), *Āryabhaṭīya of Āryabhaṭa*, Indian National Science Academy, New Delhi, 1976, p. 59.
4. Rao, S. B., *Indian Mathematics and Astronomy : Some Landmarks*, Jnana Deep Publications, Bangalore, 1994.
5. Maiti, N. L., Notes on broken bamboo problem, *Ganita Bhārati* (1994), 16, pp. 26-36.
6. Shukla and Sharma, ref. 3, p. 59; Hayashi, T., The mathematical section of the Naradapurana, *Indo-Iranian Journal* (1993), 36, pp. 1-28.
7. Gupta, R. C., Sino-Indian interaction and the great Chinese Buddhist astronomer-mathematician I- Hsing (A.D. 683-727), *Ganita Bhārati* (1989), 11, pp. 38-39.
8. Maiti, ref. 5.
9. Gupta, ref. 7.
10. *Ibid.*
11. Bag, A. K., *Mathematics in Ancient and Medieval India*, Chaukhamba Orientalia, Varanasi, 1979, p. 157.
12. Gupta, ref. 7.
13. T. L. Heath, *A Manual of Greek Mathematics*, Dover, 1963, p. 438.
14. *Ibid*, p. 51.
15. Bagheri, Mohammad, Recreational problems from *Hāsib Tabarī's Miftāḥ al-Muāmalāt*, *Ganita Bhārati* (1999), 21, pp. 1-9.
16. *Ibid.*

17. Maiti, ref. 5.
18. Bag, ref. 11, p. 157.
19. Rao, ref. 4, pp. 74-75.
20. *Ibid*, ref. 4, p. 75.
21. Bag, ref. 11, p. 158.
22. Rao, ref. 4, p. 153.
23. *Ibid*, pp. 108-109.
24. *Ibid*, pp. 166-167.
25. Mishra, V. and S. L. Singh, Height and distance problems in ancient Indian mathematics, *Ganita Bhāratī* (1996), 18, pp. 25-30.
26. Heath, ref. 13, p. 83.
27. Sarasvati, T. A., *Geometry in Ancient & Medieval India*, Motilal Banarsi das, Delhi, 1979, p. 252.
28. Sarasvati, ref. 27, pp. 254-256.
29. Singh, Parmanand, The *Ganita Kaumudi* of *Nārāyaṇa Pandita* (translation with notes), Ch. VIII, *Ganita Bhāratī* (2000), 22, p. 28.
30. Boyer, C. B., *A History of Mathematics* (revised by Uta C. Merzbach), John Wiley & Sons, New York, 1989, p. 195.
31. Gupta, ref. 7.
32. Maiti, ref. 5.
33. Bag, A. K., Ritual geometry in India and its parallelism in other cultural areas, *Indian Journal of History of Science* (1990), 25, pp. 4-19; Van der Waerden, B. L., on the Pre-Babylonian mathematics I, *Archive for the History of Exact Sciences* 1980, 23, pp. 1-25 (cf. p. 3); Maiti, ref. 5.

REFERENCES (SECTION B)

1. J. B. Fraleigh : *A First Course in Abstract Algebra*, Addison-Wesley Publishing Company (USA), 1982.
2. G. E. Martin : *The Foundations of Geometry and the Non-Euclidean Plane*, Springer-Verlag, 1998.
3. Vinod Mishra and S. L. Singh : Incircumscribing Triangles and Cyclic Quadrilaterals in Ancient and Medieval Indian Geometry, *Sugakushi Kenku* (Japan), No. 161, (1998), 1-11.
4. John Stillwell : *Elements of Algebra*, Springer -Verlag, New York, 1994.

A COMMON FIXED POINT THEOREM IN FUZZY METRIC SPACES

R. C. DIMRI* AND V. B. CHANDOLA*

(Received 19.11.2003)

ABSTRACT

Using notions of compatibility and R-weak commutativity, we prove a common fixed point theorem for six mappings in fuzzy metric spaces. Our result generalizes and fuzzifies some of the recent results in metric and fuzzy metric spaces. A related example is also furnished.

AMS subject Classification (2000) : 47H10, 54H25, 54E70.

Keywords : Fuzzy metric space, compatibility, R-weakly commuting, common fixed point.

INTRODUCTION

The theory of fuzzy sets was introduced by L. Zadeh [25] in 1965. Kramosil and Michalek [18] introduced the fuzzy metric space by generalizing the concept of probabilistic metric space to fuzzy situation. Deng [5], Erceg [6] and Kaleva and Seikkala [17] have defined fuzzy metric spaces in different ways. George and Veeramani [8] modified the concept of fuzzy metric space introduced by Kramosil and Michalek [18].

Grabiec [9] proved the contraction principle in the setting of fuzzy metric spaces introduced in [18]. For fixed point theorems in fuzzy metric spaces some of the interesting references are [1-4], [7], [9-11], [13], [14], [19], [21,22,24] etc.

Jungck [15, 16] gave the more generalized concept of compatibility than commutativity and weak commutativity in metric spaces. Mishra, Sharma and Singh [19] obtained common fixed point theorems for compatible maps on fuzzy metric spaces. Recently, Vasuki [24] introduced the concept of R-weak commutativity of

mappings in fuzzy metric space. For an excellent comparision of weaker forms of commuting maps, one may refere to Singh and Tomar [23].

In the present paper, using notions of compatibility and R-weak commutativity, we establish a common fixed point theorem for six mappings in fuzzy metric spaces in the sense of Kramosil and Michalek. Our result unifies, generalizes and fuzzifies some recent results due to Sharma [21], Singh and Chauhan [22] and Imdad and Khan [12] with less restrictive conditions on mappings.

PRELIMINARIES

Definition 1 [20]. A binary operation $* : [0, 1] \times [0, 1] \rightarrow [0, 1]$ is a continuous t-norm if $([0, 1], *)$ is a topological monoid with unit 1 such that $a * b \leq c * d$ whenever $a \leq c$ and $b \leq d$ for all $a, b, c, d \in [0, 1]$.

Examples of t-norm are $a * b = ab$ and $a * b = \min \{a, b\}$.

Definition 2 [18]. The triplet $(X, M, *)$ is a fuzzy metric space if X is an arbitrary set, $*$ is a continuous t-norm, M is a fuzzy set in $X^2 \times [0, 1]$ satisfying the following conditions :

$$(2.1) M(x, y, 0) = 0,$$

$$(2.2) M(x, y, t) = 1 \text{ for all } t > 0 \text{ iff } x = y,$$

$$(2.3) M(x, y, t) = M(y, x, t) \neq 0 \text{ for } t \neq 0,$$

$$(2.4) M(x, y, t) * M(y, z, s) \leq M(x, z, t+s),$$

$$(2.5) M(x, y, .) : [0, 1] \rightarrow [0, 1] \text{ is left continuous for all } x, y, z \in X \text{ and } s, t > 0,$$

$$(2.6) \lim_{t \rightarrow \infty} M(x, y, t) = 1 \text{ for all } x, y \text{ in } X.$$

Lemma 1 [9]. For all $x, y \in X$, $M(x, y, *)$ is nondecreasing.

Definition 3 [9]. A sequence $\{x_n\}$ in a fuzzy metric space $(X, M, *)$ is a Cauchy sequence if and only if for each $\varepsilon > 0$, $t > 0$, there exists $n_0 \in \mathbb{N}$ such that

$$M(x_n, x_m, t) > 1 - \varepsilon \text{ for all } n, m \geq n_0.$$

Definition 4 [9]. A sequence $\{x_n\}$ in a fuzzy metric space $(X, M, *)$ is said to converge to x if and only if for each $\varepsilon > 0$, $t \geq 0$, there exists $n_0 \in \mathbb{N}$ such that

$$M(x_n, x, t) > 1 - \varepsilon \text{ for all } n \geq n_0.$$

Definition 5 [22]. Self mappings F and G of a fuzzy metric space $(X, M, *)$ are said to be compatible if and only if $M(FGx_n, GFx_n, t) \rightarrow 1$ for all $t > 0$, whenever $\{x_n\}$ is a sequence in X such that $Fx_n, Gx_n \rightarrow y$ for some y in X .

Definition 6 [24]. The mappings F and G of a fuzzy metric space $(X, M, *)$ into itself are R -weakly commuting provided there exists some positive real number R such that

$$M(FGx, GFx, tR) \geq M(Fx, Gx, t) \text{ for all } x \text{ in } X.$$

Obviously, weak commutativity implies R -weak commutativity. However, R -weak commutativity implies weak commutativity only when $R \leq 1$.

Lemma 2 [3]. Let $\{x_n\}$ be a sequence in a fuzzy metric space $(X, M, *)$. If there exists a number $k \in (0, 1)$ such that

$M(x_{n+2}, x_{n+1}, kt) \geq M(x_{n+1}, x_n, t)$ for all $t > 0$ and $n = 1, 2, \dots$, then $\{x_n\}$ is a Cauchy sequence in X .

Lemma 3 [19]. If for all $x, y \in X$, $t > 0$ and for a number $k \in (0, 1)$,

$$M(x, y, kt) \geq M(x, y, t) \text{ then } x = y.$$

MAIN RESULTS

Theorem 1. Let A, B, S, T, I and J be self mappings of a complete fuzzy metric space $(X, M, *)$ with continuous t -norm defined by $a * b = \min \{a, b\}$, $a, b \in [0, 1]$, satisfying the following conditions :

(1.1) $AB(X) \subset J(X), ST(X) \subset I(X),$

(1.2) if either (a) $\{AB, I\}$ are compatible, I or AB is continuous and (ST, J) are R -weakly commuting, or

(a') $\{ST, J\}$ are compatible, J or ST is continuous and (AB, I) are R -weakly commuting,

(1.3) there exists a number $k \in (0, 1)$ such that

$$M(ABx, STy, kt) \geq \min \{M(ABx, Jy, 2t), M(STy, Ix, t), M(ABx, Ix, t),$$

$$M(STy, Jy, t), M(Ix, Jy, t)\} \text{ for all } x, y \text{ in } X \text{ and } t > 0.$$

(1.4) $\lim M(x, y, t) \rightarrow 1$ as $t \rightarrow \infty$, for all x, y in X ; then AB, ST, I and J have a unique common fixed point. Furthermore, if the pairs $(A, B), (A, I), (B, I), (S, T), (S, J)$ and (T, J) are commuting mappings then A, B, S, T, I and J have a unique common fixed point.

Proof. Let x_0 be an arbitrary point in X . Since $AB(X) \subset J(X)$, we can find a point x_1 in X such that $ABx_0 = Jx_1$. Also since $ST(X) \subset I(X)$ we can choose a point x_2 with $STx_1 = Ix_2$. Using this argument repeatedly one can construct a sequence $\{y_n\}$ such that

$$y_{2n} = ABx_{2n} = Jx_{2n+1}, y_{2n+1} = STx_{2n+1} = Ix_{2n+2}$$

for $n = 0, 1, 2, \dots$

From (1.3)

$$\begin{aligned} M(y_{2n+1}, y_{2n+2}, kt) &= M(STx_{2n+1}, ABx_{2n+2}, kt) \\ &\geq \min \{M(ABx_{2n+2}, Jx_{2n+1}, 2t), M(STx_{2n+1}, Ix_{2n+2}, t), \\ &\quad M(ABx_{2n+2}, Ix_{2n+2}, t), M(STx_{2n+1}, Jx_{2n+1}, t), \\ &\quad M(Ix_{2n+2}, Jx_{2n+1}, t)\} \\ &= \min \{M(y_{2n+2}, y_{2n}, 2t), M(y_{2n+1}, y_{2n+1}, t), \\ &\quad M(y_{2n+2}, y_{2n+1}, t), M(y_{2n+1}, y_{2n}, t), M(y_{2n+1}, y_{2n}, t)\} \\ &= \min \{M(y_{2n}, y_{2n+1}, t), M(y_{2n}, y_{2n+2}, 2t), 1\}, \end{aligned}$$

$$\geq \min \{M(y_{2n}, y_{2n+1}, t), M(y_{2n}, y_{2n+1}, t), M(y_{2n+1}, y_{2n+2}, t), \\ M(y_{2n}, y_{2n+1}, t)\}.$$

In general

$$M(y_n, y_{n+1}, kt) \geq M(y_{n-1}, y_n, t).$$

By Lemma 2, $\{y_n\}$ is a Cauchy sequence in X . Since X is complete, $\{y_n\}$ converges to some point z in X . Thus $\{ABx_{2n+1}\}$, $\{Jx_{2n+1}\}$, $\{STx_{2n+1}\}$ and $\{Ix_{2n+2}\}$ also converge to the point z .

Case I : Let us now assume that I is continuous. Then the sequences $\{I^2x_{2n}\}$ and $\{IABx_{2n}\}$ converge to Iz . Thus for $t > 0$, $\lambda \in (0,1)$, there exists an integer $n_0 \in \mathbb{N}$ such that

$$M(IABx_{2n}, Iz, t/2) > 1 - \lambda, \quad M(I^2x_{2n}, Iz, t/2) > 1 - \lambda \text{ for all } n \geq n_0. \quad (1.5)$$

As $\{AB, I\}$ are compatible, using (1.5) we get,

$$M(ABIx_{2n}, Iz, t) \geq \min \{M(ABx_{2n}, IABx_{2n}, t/2), M(IABx_{2n}, Iz, t/2)\} \\ > 1 - \lambda \text{ for all } n \geq n_0.$$

Hence $ABIx_{2n} \rightarrow Iz$. (1.6)

Now using (1.3),

$$M(ABIx_{2n}, STx_{2n+1}, kt) \geq \min \{M(ABIx_{2n}, Jx_{2n+1}, 2t), M(STx_{2n+1}, I^2x_{2n}, t), \\ M(ABIx_{2n}, I^2x_{2n}, t), M(STx_{2n+1}, Jx_{2n+1}, t), M(I^2x_{2n}, Jx_{2n+1}, t)\}.$$

On letting $n \rightarrow \infty$ and using (1.6), we get

$$M(Iz, z, kt) \geq M(Iz, z, t).$$

So by Lemma 2, $Iz = z$. (1.7)

Again using (1.3),

$$M(ABz, STx_{2n+1}, kt) \geq \min \{M(ABz, Jx_{2n+1}, 2t), M(STx_{2n+1}, Iz, t), M(ABz, Iz, t), \\ M(STx_{2n+1}, Jx_{2n+1}, t), M(Iz, Jx_{2n+1}, t)\}.$$

Letting $n \rightarrow \infty$ and using (1.7),

$$M(ABz, z, 2t) \geq \min \{M(ABz, z, 2t), M(z, Iz, t), M(ABz, Iz, t), M(z, z, t), M(Iz, z, t)\} \\ = M(ABz, z, t).$$

Therefore $M(ABz, z, kt) \geq M(ABz, z, t)$. By Lemma 3, this implies that

$$ABz = Iz = z. \quad (1.8)$$

Since $AB(X) \subset J(X)$, there always exists a point z' such that

$$J(z') = AB(z) = z \text{ so that } STz = ST(Jz').$$

Now

$$M(z, STz', kt) = M(ABz, STz', kt) \\ \geq \min \{M(ABz, Jz', 2t), M(STz', Iz, t), M(ABz, Iz, t), \\ M(STz', Jz', t), M(Iz, Jz', t)\} \\ \geq M(STz', z, t).$$

$$\text{Hence by Lemma 3 } STz' = z = Jz', \quad (1.9)$$

which shows that z' is the coincidence point of ST and J . Now using the R -weak commutativity of (ST, J) and from (1.9) we have,

$$M(STJz', JSTz', Rt) \geq M(STz', Jz', t) \rightarrow 1.$$

Therefore $ST(Jz') = J(STz')$.

$$\text{Hence } STz = ST(Jz') = JSTz' = Jz, \quad (1.10)$$

which shows that z is also coincidence point of the pair (ST, J) .

Using (1.3), (1.7), (1.8) and (1.10),

$$\begin{aligned}
 M(z, STz, kt) &= M(ABz, STz, kt) \\
 &\geq \min \{M(ABz, Jz, 2t), M(STz, Iz, t), M(ABz, Iz, t), \\
 &\quad M(STz, Jz, t), M(Iz, Jz, t)\} \\
 &\geq M(z, STz, t),
 \end{aligned}$$

giving $z = STz = Tz$ by using Lemma 3.

Therefore z is a common fixed point of AB , I , ST and J .

Case II: Now suppose that AB is continuous so that the sequences $\{(AB)^2 x_{2n}\}$ and $\{ABIx_{2n}\}$ converge to ABz . Since (AB, I) are compatible, therefore

$$\begin{aligned}
 M(IABx_{2n}, ABz, t) &\geq \{M(IABx_{2n}, ABIx_{2n}, t/2), M(ABIx_{2n}, ABz, t/2)\} \\
 &> 1 - \lambda \text{ for all } n \geq n_0
 \end{aligned}$$

Hence $IABx_{2n} \rightarrow ABz$. (1.11)

Using (1.3), (1.11) and letting $n \rightarrow \infty$, we get

$$M(ABz, z, kt) \geq M(AB, z, t).$$

Therefore by Lemma 3, we have $ABz = z$.

As earlier there exists z' in X such that $ABz = z = Jz'$. Then

$$\begin{aligned}
 M((AB)^2 x_{2n}, STz', kt) &\geq \min \{M(AB)^2 x_{2n}, Jz', 2t), M(STz', IABx_{2n}, t), \\
 &\quad M((AB)^2 x_{2n}, IABx_{2n}, t), M(STz', Jz', t), \\
 &\quad M(IABx_{2n}, Jz', t)\}
 \end{aligned}$$

which on taking the limit $n \rightarrow \infty$ reduces to

$$M(z, STz', kt) \geq M(z, STz', t).$$

By Lemma 3, this implies that $STz' = z = Jz'$.

Thus z' is the coincidence point of (ST, J) . Since the pair (ST, J) are R -weakly

commuting, then $STz = Jz$.

$$\text{Further } M(ABx_{2n}, STz, kt) \geq \min \{M(ABx_{2n}, Jz, 2t), M(STz, Ix_{2n}, t), M(ABx_{2n}, Ix_{2n}, t), \\ M(STz, Jz, t), M(Ix_{2n}, Jz, t)\}$$

which on letting $n \rightarrow \infty$ reduces to

$$M(z, STz, kt) \geq M(z, STz, t) \text{ gives}$$

$$STz = z = Jz.$$

Since $ST(X) \subset I(X)$, there exists a point z'' in X such that

$$Iz'' = STz = z, \text{ then}$$

$$\begin{aligned} M(ABz'', z, kt) &= M(ABz'', STz, kt) \\ &\geq \min \{M(ABz'', Jz, 2t), M(STz, Iz'', t), M(ABz'', Iz'', t), \\ &\quad M(STz, Jz, t), M(Iz'', Jz, t)\} \\ &\geq M(ABz'', z, t) \end{aligned}$$

which gives $ABz'' = z$.

Also (AB, I) are compatible and hence weakly commuting, we obtain,

$$\begin{aligned} M(ABz, Iz, t) &= M(ABz'', IABz'', t) \geq M(ABz'', Iz, t) \\ &= M(z, z, t) \rightarrow 1. \end{aligned}$$

Therefore $ABz = Iz = z$.

Thus z is a common fixed point of AB, ST, I and J .

If the mappings ST or J is continuous instead of AB or I , then the proof that z is a common fixed point of AB, ST, I and J is similar.

Let w be another fixed point of I, J, AB and ST then

$$\begin{aligned}
 M(z, w, kt) &= M(ABz, STw, kt) \\
 &\geq \min \{M(ABz, Jw, 2t), M(STw, Iz, t), M(ABz, Iz, t), \\
 &\quad M(STw, Jw, t), M(Iz, Jw, t)\}
 \end{aligned}$$

which gives $M(z, w, kt) \geq M(z, w, t)$

Hence $z = w$.

Finally, we need to prove that z is also a common fixed point of A, B, S, T, I and J .

Using commutativity of (A, B) and (S, T) , it is easy to prove that

$$Az = z = Bz \text{ and } Sz = z = Tz.$$

$$Now, M(Az, Sz, kt) = M(A(ABz), S(STz), kt)$$

$$= M(A(BAz), S(TSz), kt)$$

$$= M(AB(Az), ST(Sz), kt)$$

$$\geq \min \{M(AB(Az), JSz, 2t), M(STSz, IAz, t), M(ABAz, IAz, t),$$

$$M(STSz, JSz, t), M(IAz, JSz, t)\},$$

Using commutativity of $(A, B), (S, T), (I, A)$ and (J, S) we have

$M(Az, Sz, kt) \rightarrow 1$ so $Az = Sz$. Similarly it can be shown that $Bz = Tz$. Thus z is the unique common fixed point of A, B, S, T, I and J .

Remark 1. Theorem 1 remains true if we replace $a * b$ by $a.b$.

If we put $I = J, A = S$ and $B = T =$ Identity map in Theorem 1, we get the following result.

Corollary. Let A and J be self maps of a complete fuzzy metric space $(X, M, *)$ with continuous t -norm defined by $a * b = \min(a, b)$, $a, b \in (0, 1)$ such that (A, J) are compatible, J is continuous and

$$M(Ax, Ay, kt) \geq \min \{M(Ax, Jy, 2t), M(Ay, Jx, t), M(Ax, Jx, t), M(Ay, Jy, t)\}$$

$M(Jx, Jy, t)\}$

for all x, y in X , $t > 0$ and $k \in (0, 1)$.

If for all x, y in X , $\lim (x, y, t) \rightarrow 1$ as $t \rightarrow \infty$, then A and J have a unique common fixed point.

Remark 2. A number of fixed point theorems may be obtained for two to four mappings in metric, probabilistic and fuzzy metric spaces as the special cases from Theorem 3.1.

Example. Let $X = [0, 1]$. For $a, b \in [0, 1]$ define $a * b = \min(a, b)$ and

$M(x, y, t) = [\exp(|x-y|)/t]^{-1}$ for all $x, y \in X$ and $t > 0$. Then clearly $(X, M, *)$ is a fuzzy metric space (see example 2.7 in [8]).

Define mappings A, B, S, T, I and J as

$Ax = x/64, Bx = x/16, Sx = x/32, Tx = x/8, Ix = x/4, Jx = x$ for all $x \in [0, 1]$.

Then $AB(X) = [0, 1/1024] \subset [0, 1] = J(X)$ and $ST(X) = [0, 1/256] \subset [0, 1/4] = I(X)$.

If $\lim x_n = 0$ where $\{x_n\}$ is a sequence in X , then clearly (AB, I) are compatible.

Also (ST, J) are R-weakly commuting.

Now if we take $1/256 \leq k \leq 1$, $t > 0$, $x, y \in [0, 1]$ then condition (1.3) of Theorem 1 is also satisfied. Here zero is the unique common fixed point of A, B, S, T, I and J .

REFERENCES

1. R. Badard: Fixed point theorems for fuzzy numbers, *Fuzzy Sets and Systems* 13(1984) 296-302.
2. S.-S. Chang, Y.J. Cho, B.S. Lee, J.S. Jung and S.M. Kang: Coincidence point and minimization theorems in fuzzy metric spaces, *Fuzzy Sets and Systems* 88(1) (1997) 119-128.

3. Y. J. Cho: Fixed points in fuzzy metric spaces, *J. Fuzzy Math.* 5(4)(1997) 949-962.
4. R. Chugh and S. Kumar : Common fixed point theorem in fuzzy metric spaces, *Bull. Cal. Math. Soc.* 94 (1) (2002) 17-22.
5. Z. K. Deng: Fuzzy pseudo-metric space, *J. Math. Anal. Appl.* 86(1982) 74-95.
6. M. A. Erceg, Metric space in fuzzy set theory, *J. Math. Anal. Appl.* 69(1979) 205-230.
7. J. X. Fang: On fixed point theorems in fuzzy metric spaces, *Fuzzy Sets and Systems* 46(1992) 107- 113.
8. A. George and P. Veeramani: On some results in fuzzy metric spaces, *Fuzzy Sets and Systems* 64(1994) 395-399.
9. M. Grabiec: Fixed points in fuzzy metric spaces, *Fuzzy Sets and Systems* 27(1988) 385-389.
10. O. Hadžić: Fixed point theorems for multi- valued mappings in some classes of fuzzy metric spaces, *Fuzzy Sets and Systems* 29(1989) 115-125.
11. O. Hadžić and Endre Pap: A fixed point theorem for multi-valued mappings in probabilistic metric spaces and an application in fuzzy metric spaces, *Fuzzy Sets and Systems* 127(2002) 333- 344.
12. M. Imdad and Q.H. Khan: A common fixed point theorem for six mappings satisfying a rational inequality, *Ind. J. Math.* 44 (1) (2002) 47-57.
13. J. S. Jung, Y. J. Cho, S.-S. Chang and S.M. Kang: Coincidence theorems for set valued mappings and Ekeland's variational principle in fuzzy metric spaces, *Fuzzy Sets and Systems* 79 (1996) 239-250.
14. J. S. Jung, Y. J. Cho and J. K. Kim: Minimization theorems for fixed point

theorems in fuzzy metric spaces and applications, *Fuzzy Sets and Systems* 61(1994) 199-207.

15. G. Jungck: Compatible mappings and common fixed points, *Internat. J. Math. Math. Sci.* 9 (1986) 779-791.
16. G. Jungck: Compatible mappings and common fixed point (2), *Internat. J. Math. Math. Sci.* 11(2) (1988) 285-288.
17. O. Kaleva and S. Seikkala, On fuzzy metric spaces, *Fuzzy Sets and Systems* 12(1984) 215-229.
18. I. Kramosil and J. Michalek: Fuzzy metric and Statistical metric spaces, *Kybernetika* 11(1975) 326- 334.
19. S. N. Mishra, N. Sharma and S.L. Singh: Common fixed point of maps on fuzzy metric spaces, *Internet J. Math. Math. Sci* 17(1994) 253-258.
20. B. Schweizer and A. Sklar: Statistical metric spaces, *Pacific J. Math.* 10 (1960) 313-334.
21. S. Sharma: Common fixed point theorems in fuzzy metric spaces, *Fuzzy Sets and Systems*, 127(2002) 345-352.
22. B. Singh and M.S. Chauhan: Common fixed points of compatible maps in fuzzy metric spaces, *Fuzzy Sets and Systems*, 115(2000) 471- 475.
23. S. L. Singh and Anita Tomar: Weaker forms of commuting maps and existence of fixed points, *J. Korea Soc. Math. Educ. Ser. B: Pure Appl. Math.* 10(3) (2003), 145-161.
24. R. Vasuki: Common fixed points for R-weakly commuting maps in fuzzy metric spaces, *Indian J. Pure Appl. Math.* 30(4) (1999) 419-423.
25. L . A. Zadeh: Fuzzy Sets, *Inform. Control* 8 (1965) 338-353.

KINETICS AND MECHANISM OF Mn (II) CATALYSED PERIODATE OXIDATION OF O-TOLUIDINE

R.D. KAUSHIK*, SUREKHA KANNAUJIA* AND SHASHI*

(Received 29.09.2003 and in revised form 24.12.2003)

ABSTRACT

Kinetic-mechanistic studies made for periodate oxidation of o-toluidine (OMA) in acetone-water medium, have been used for derivation of rate law and proposing mechanism which satisfy various observations like first order in each reactant and catalyst, stoichiometry (1 mol OMA : 2 mol periodate), thermodynamic parameters, main product identified (Methyl-1,4 - benzoquinone), the Michaelis-Menten type kinetics being followed with respect to both reactants, the rate - pH profile, increase in rate with the increase in dielectric constant of the medium, negative primary linear type effect of ionic strength and no effect of free radical scavengers on the reaction rate.

Key words and phrases: Kinetics and mechanism, Periodate oxidation, o-toluidine, Mn (II) catalyst, Methyl-1, 4- benzoquinone.

INTRODUCTION

Reports on kinetic- mechanistic studies on the Mn (II) catalysed periodate oxidation of aromatic amines are very few [3-8] In continuation to our earlier communications on kinetic-mechanistic studies on uncatalysed periodate oxidation of few aromatic amines [9-14] and Mn (II) catalysed oxidation of 4 - chloro-2-methylaniline [15], we are reporting the kinetic-mechanistic studies on Mn (II) catalysed periodate oxidation of o-toluidine in acetone-water medium in present paper.

MATERIALS AND METHODS

Chemicals of E. Merck/CDH A.R. grade were used after distillation/recrystallization. Triply distilled water was used for preparation of the solutions. The progress of the reaction was followed spectrophotometrically by recording the absorbance at 490 nm, the λ_{max} of reaction mixture in the duration in which the λ_{max} did not change. The pH was maintained at 4.5 by using Thiel, Schultz and Koch buffer [1] in all kinetic runs except when the effect of pH was studied.

* Department of Chemistry, Gurukul Kangri University, Haridwar - 249404 (India).

NaCl solutions were used for maintaining the ionic strength (μ) in the kinetic runs. Plain mirror method and Guggenheim's method were used for evaluation of initial rates $[(dA/dt)_0]$ and pseudo first order rate constant k_1 (or second order rate constant k_2) respectively.

RESULTS AND DISCUSSION

Initially the reaction mixture was yellow in color, which changed into orange-wine color in about 2 minutes. On standing overnight, it changed to brown color followed by precipitation. The reaction mixture was prepared by taking periodate in excess and kept overnight. After precipitation, the supernatant liquid was extracted with petroleum ether (40-60°C). On evaporation of the solvent, a yellow colored compound with melting point 68°C was obtained. This compound responded positively to the test for quinone.

The UV-VIS spectrum of this compound in ethanol showed λ_{max} at 240 and 315 nm. The melting point and λ_{max} are in good agreement with the values reported for methyl-1,4-benzoquinone [2, 18]. The I.R. spectrum of this compound in KBr showed the presence of bands at 3054 cm^{-1} (s) (due to ring C-H stretch), 2925 cm^{-1} (s), 2854 cm^{-1} (s) (due to C-H stretching vibrations of CH_3), 1709 cm^{-1} (w) & 1744 cm^{-1} (w) (due to overtones and combination bands), 1632 cm^{-1} (s) (indicating the presence of C = O on 1, 4-benzoquinone pattern with the possibility that the position of this band gets lowered due to + I effect of alkyl group [19]), 1514 & 1457 cm^{-1} (s) (due to C = C ring stretch), 1330 & 1135 cm^{-1} (m) (due to in-plane C - H bending in the ring), 751 & 664 cm^{-1} (s) (due to out-of-plane C = C and = C - H bending modes and substitution pattern in the ring). NMR spectrum in CDCl_3 showed the peaks at $\delta = 7.091$, D, 2H; at $\delta = 6.658$, S, 1H; and at $\delta = 2.156$, S, 3H. The singlet at $\delta = 2.156$ may be due to protons of $-\text{CH}_3$ group while other singals are due to 3 protons of the ring [19].

On the basis of melting point [Literature: M.P. = 69°C, yellow plates or needles] [2] and spectral studies [Literature: UV-VIS λ_{max} at 246 nm ($\log \epsilon = 4.14$), at 312 nm ($\log \epsilon = 2.77$), at 429 nm ($\log \epsilon = 1.28$) and at 438 nm ($\log \epsilon = 1.27$)], the characteristic I.R. peaks and NMR signals matching the structure of methyl-1,4-benzoquinone [18-19], this compound may be methyl-1,4-benzoquinone which is the main product of the reaction under consideration.

1 mol CMA consumed 2 moles of periodate as determined iodometrically. The data (table-1) indicated 2nd order for the reaction, being first order in each reactant. Linear relation between concentration of the reactants and rate supported the 2nd order kinetics. In pseudo first order conditions (table-2), the $[(dA/dt)]^{-1}$ or k_1^{-1} vs $[S]^{-1}$ plots were linear with almost negligible intercept, suggesting the Michaelis-Menten type kinetics being followed with respect to both reactants with the possibility of formation of a fast decaying intermediate complex between reactants [15-16].

Data in table-3 established the 1st order in catalyst. Rate-pH profile showed a maxima at pH = 4.5 (table-4, Fig.-1). A linear relation between $\log (dA/dt)$, or $\log k_2$ and $1/D$ with negative slope (where D is the dielectric constant of the medium) and a primary linear type plot between $\log (dA/dt)$, or $\log k_2$ vs ionic strength (μ) that were obtained from the data in table-5, indicated an ion-dipole interaction in this reaction. This is supported by our observation that free radical scavengers like acryl amide and allyl alcohol exerted no effect on reaction rate.

Arrhenius plot was made between $30 \pm 0.1^\circ\text{C}$ to $45 \pm 0.1^\circ\text{C}$ and the values of different thermodynamic parameters evaluated taking $[\text{OMA}] = 5.0 \times 10^{-4} \text{ M}$, $[\text{NaIO}_4] = 5.0 \times 10^{-3} \text{ M}$, $[\text{Mn}^{++}] = 4.0 \times 10^{-6} \text{ M}$, pH = 4.5 and acetone

= 10.0%
(v/v), are $E_a = 8.10 \text{ k.cal mol}^{-1}$; $A = 4.88 \times 10^3 \text{ lit. mol}^{-1} \text{ sec}^{-1}$; $\Delta S^\ddagger = -60.17 \text{ e. u.}$; $\Delta F^\ddagger = 18.35 \text{ k. cal. mol}^{-1}$ and $\Delta H^\ddagger = 7.49 \text{ k.cal. mol}^{-1}$. Low value of energy of activation and high frequency factor are characteristic of a bimolecular reaction in the solution in which the reacting species are large. A large negative value of ΔS^\ddagger suggests the formation of strongly solvated, charged and rigid transition state.

Table-1. Determination of order w.r. t. reactants.

$\lambda_{\text{max}} = 475 \text{ nm}$; pH = 4.5; Acetone* = 7.5%; Acetone** = 10.0% (v/v); Temp = $35 \pm 0.1^\circ\text{C}$; $[\text{Mn}^{++}] = 8.0 \times 10^{-6} \text{ M}$.

$[\text{OMA}] \times 10^3 \text{ M}$	1.0*	2.0*	3.0*	5.0*	6.0*	1.0**	1.0**	1.0**	1.0**	1.0**
$[\text{NaIO}_4] \times 10^3 \text{ M}$	1.0	1.0	1.0	1.0	1.0	2.0	3.0	4.0	6.0	7.0
$[(dA/dt)] \times 10^2 \text{ (min}^{-1}\text{)}$	4.8	9.7	14.3	23.8	28.7	7.4	10.8	14.3	21.5	25.2

Table-2. Variation of OMA and periodate

$\lambda_{\max} = 475$ nm; pH = 4.5; Acetone = 10.0% (v/v); Temp = $35.0 \pm 0.1^\circ\text{C}$;

$[\text{Mn}^{++}] = 4.0 \times 10^{-6}$ M; $[\text{Mn}^{++}]^{**} = 8.0 \times 10^{-6}$ M.

$[\text{OMA}] \times 10^3$ M	10.0*	12.0*	14.0*	18.0*	0.5**	0.5**	0.5**	0.5**	0.5**
$[\text{NaIO}_4] \times 10^3$ M	1.0	1.0	1.0	1.0	5.0	7.0	8.0	9.0	10.0
$(dA/dt)_i \times 10^2$ (min ⁻¹)	4.8	5.8	6.9	8.7	3.8	4.4	4.8	5.6	6.3
$k, \times 10^3$ (sec ⁻¹)	3.84	4.60	5.37	6.91	4.03	5.57	6.14	6.91	7.68
$k, \times 10^2$ (L.mol ⁻¹ . sec ⁻¹)	3.84	3.83	3.83	3.84	8.06	7.95	7.68	7.68	7.68
Average $k,$	38.36×10^{-2}				78.08×10^{-2}				

Table-3. Determination of order w. r. t. Mn^{++} .

$\lambda_{\max} = 475$ nm; pH = 4.5; $[\text{NaIO}_4] = 5.0 \times 10^{-3}$ M; $[\text{OMA}] = 5.0 \times 10^{-4}$ M; Temp. = $35.0 \pm 0.1^\circ\text{C}$; Acetone = 15.0% (v/v).

$[\text{Mn}^{++}] \times 10^6$ M	4.0	5.0	6.0	7.0	8.0	9.0	10.0
$(dA/dt)_i \times 10^2$ (min ⁻¹)	2.6	3.2	3.9	4.6	5.2	5.9	6.5

Table-4. Effect of pH on reaction rate.

$\lambda_{\max} = 475$ nm; $[\text{OMA}] = 1.0 \times 10^{-3}$ M; $[\text{NaIO}_4] = 1.0 \times 10^{-2}$ M; Acetone = 10.0% (v/v); Temp. = $35.0 \pm 0.1^\circ\text{C}$; $[\text{Mn}^{++}] = 4.0 \times 10^{-6}$ M

pH	3.0	3.5	4.0	4.5	5.0	5.5	6.0	6.5	7.0	7.5	8.0	8.5	9.0
$(dA/dt)_i \times 10^{-3}$ (min ⁻¹)	0.72	1.92	50.0	86.0	52.0	34.0	20.0	7.6	4.4	2.8	2.6	1.1	0.71

Table-5. Effect of Dielectric constant of medium (D) and ionic strength (μ) on the reaction rate

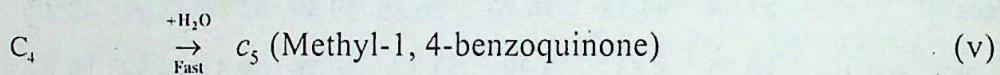
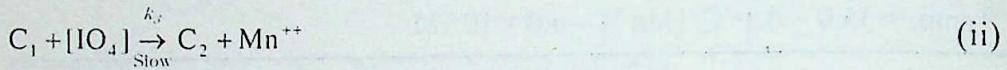
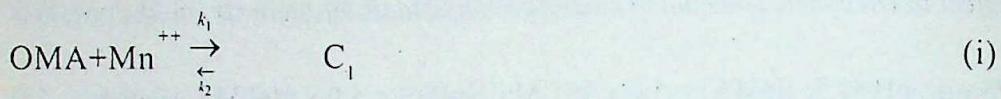
$\lambda_{\text{max}} = 475 \text{ nm}$; pH = 4.5; $[\text{OMA}] = 5.0 \times 10^{-4} \text{ M}$; $[\text{NaIO}_4] = 5.0 \times 10^{-3} \text{ M}$; * Acetone = 5.0 % (v/v); Temp. = $35.0 \pm 0.1^\circ \text{C}$; $[\text{Mn}^{++}] = 4.0 \times 10^{-6} \text{ M}$

D	71.0	69.7	68.4	66.8	--	--	--	--
* $\mu \times 10^3$	--	--	--	--	12.0	18.0	26.0	34.0
$(dA/dt)_1 \times 10^3 \text{ (min}^{-1})$	44.0	36.0	31.0	26.0	74.0	59.0	45.0	34.0
$k_1 \times 10^4 \text{ (sec}^{-1})$	53.74	44.14	36.46	26.86	59.49	53.74	47.98	42.22
$k_2 \times 10^4 \text{ (L. mol}^{-1} \text{. sec}^{-1})$	10.75	8.83	7.29	5.37	20.76	20.31	19.82	19.27

The increase in the rate from pH 3.0 to 4.5 may be due to the decrease in the protonation of OMA from pH 3.0 to 4.5 (table-4, Fig.-1), which makes greater concentrations of OMA available for the reaction. This leads to the assumption that unprotonated OMA is the reactive species in present case. Second part of this profile suggests that out of various species in which periodate exists, periodate monoanion $[\text{IO}_4^-]$ is the reactive species in present investigation [17]. Pavolva et al. [17] established that $[\text{IO}_4^-]$ goes on decreasing with increase in pH beyond the value 4.5 decreasing thereby the rate of reaction beyond pH 4.5. Similar rate- pH profiles in case of periodate oxidation of some other aromatic amines [9-15,20] have been reported earlier by us.

Based on these results, the proposed mechanism (chart) might involve the lone pair of electrons on nitrogen atom of OMA for the co-ordinate bond formation between OMA and Mn^{++} species in a reversible step to form complex C_1 in step (i). C_1 , in turn, interacts with IO_4^- in slow and rate determining step (ii) to give C_2 which changes by fast hydrolysis into C_3 . The formation of a charged intermediate complex C_2 by the attack of IO_4^- on the nitrogen of anilino group and stabilization of positive charge on nitrogen of this group, has already been established and supported by LFER studies for the uncatalysed periodate oxidation of few aromatic amines [21]. In addition, a high negative value of entropy of activation and the effect of dielectric constant on the reaction rate support the involvement of solvation effects in this reaction.

It should also be noted that the initial part of the reaction is significant in the present case and the second molecule of IO_4^- reacting later to give C_4 is not significant. C_4 changes by fast hydrolysis to give C_5 i.e. the main product of reaction that has been isolated, separated and characterized in this case as methyl-1, 4-benzoquinone. The overall process may be represented as follows:



On applying steady state treatment to C_1 , the rate law in terms of rate of loss of $[\text{IO}_4^-]$ may be derived as follows:

Rate of loss of $[\text{IO}_4^-]$ or $-\frac{d[\text{IO}_4^-]}{dt} = k_3 [\text{C}_1][\text{IO}_4^-]$ = Rate of loss of C_1

$$\text{or } -\frac{d[\text{C}_1]}{dt} = k_3 [\text{C}_1][\text{IO}_4^-] \quad \text{---(1)}$$

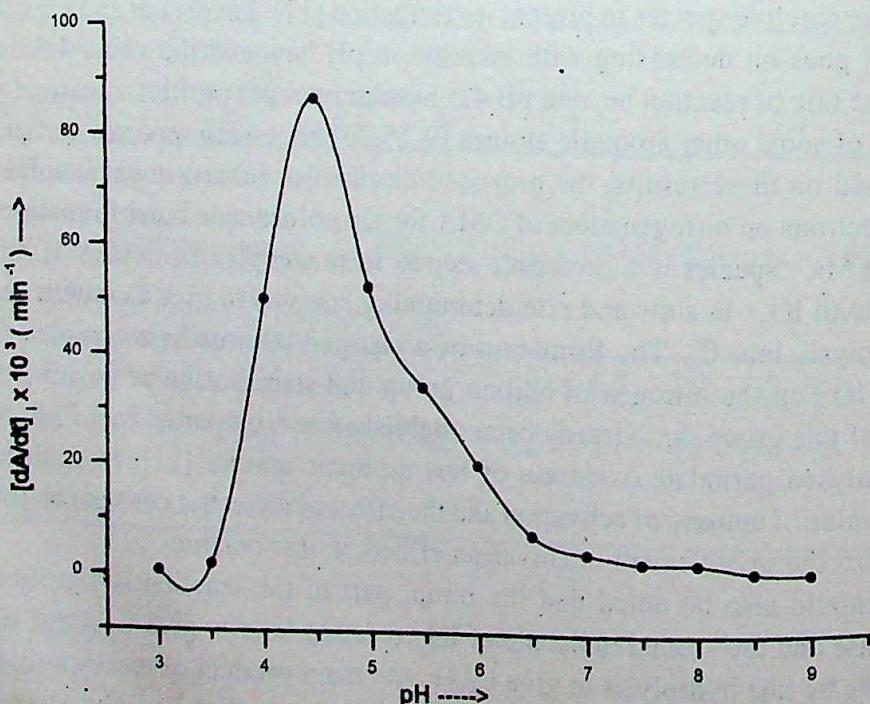
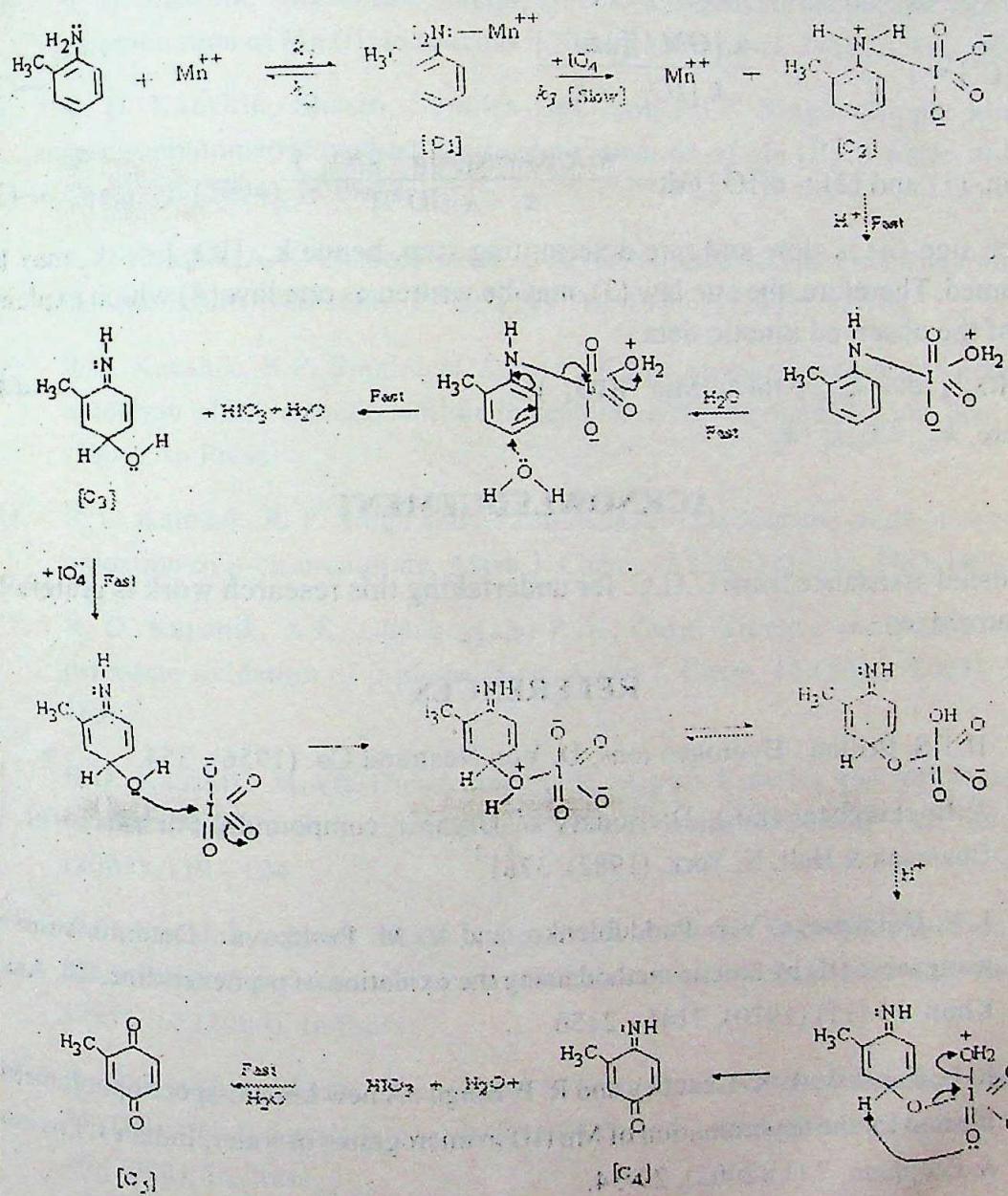


Fig. 1- Rate - pH profile

CHART - I



Methyl-1,4-benzoquinone

Rate of formation of $C_1 = + d [C_1]/dt = k_1 [OMA] [Mn^{++}] - k_2 [C_1]$

At steady state, $- d [C_1]/dt = + d [C_1]/dt$

Therefore, $k_3 [C_1] [IO_4^-] = k_1 [OMA] [Mn^{++}] - k_2 [C_1]$

Or $[C_1] \{k_3 [IO_4^-] + k_2\} = k_1 [OMA] [Mn^{++}]$

$$\text{Or } [C_1] = \frac{k_1 [OMA] [Mn^{++}]}{k_3 [IO_4^-] + k_2} \quad \text{---(2)}$$

$$\text{From, (1) and (2), } - d [IO_4^-]/dt = \frac{k_3 k_1 [OMA] [Mn^{++}] [IO_4^-]}{k_2 + k_3 [IO_4^-]} \quad \text{---(3)}$$

Since step (ii) is slow and rate determining step, hence $k_3 \cdot [IO_4^-] \ll k_2$ may be assumed. Therefore, the rate law (3), may be written as rate law (4) which explains all of the observed kinetic data.

$$- d [IO_4^-]/dt = k_{obs} \cdot [OMA] [Mn^{++}] [IO_4^-] \quad \text{---(4)}$$

where, $k_{obs} = k_3 \cdot k_1 / k_2$

ACKNOWLEDGEMENT

Financial assistance from U.G.C. for undertaking this research work is gratefully acknowledged.

REFERENCES

1. H.T.S. Britton : Hydrogen ions, D. Von Nostrand Co. (1956), 354.
2. J. Buckingham (Ed.): Dictionary of Organic compounds, 5th edn., Vol. 4, Chapman & Hall, N. York, (1982), 3761.
3. I. F. Dolmanova, V.P. Poddubienko, and V. M. Peshkova : Determination of manganese (II) by kinetic method using the oxidation of p-phenetidine, Zh. Anal. Khim. 25 (11) (1970), 2146 - 2150.
4. R. D. Kaushik, A. K. Chaubey and R. P. Singh : A new kinetic-spectrophotometric method for the determination of Mn (II) in micrograms in water, Indian J. Environ. & Ecoplann. 7 (1) (2003), 29-34.
5. R. D. Kaushik, Amrita, and R.P. Singh : Kinetic method for microgram determination of Mn (II) based on its catalytic effect on periodate oxidation of m-chloroaniline, J. Curr. Sci., 4 (1) (2004), In Press.

6. R.D. Kaushik, Sumitra Devi, Shashi and Amrita: Periodate oxidation of 2,3 - dimethylaniline---Determination of Mn (II) in nanograms in aqueous medium, Indian J. Environ. & Ecopl., 8 (1) (2004), In Press.
7. R. D. Kaushik, Amrita and Sumitra Devi: An improved method for nanogram determination of Mn (II) in aqueous medium, J. Curr. Sci. 3 (1) (2003), 197-202.
8. R. D. Kaushik, Shashi, Sumitra Devi and R. P. Singh: Simple kinetic-spectrophotometric method for the determination of Mn (II) in water, Asian J. Chem. 16 (2004), In Press.
9. R. D. Kaushik, A. K. Chaubey and R. K. Arya: Periodate oxidation of o-chloroaniline-akinetic-mechanistic study, J. Nat. Phys. Sci. 16 (1-2) (2002), 97-104.
10. R.D. Kaushik, R.P. Singh and Amrita : Kinetics and mechanism of periodate oxidation of 2,3-dimethylaniline in acetone- water medium, J. Curr. Sci., 4 (1) (2004), In Press.
11. R. D. Kaushik, R. P. Singh and Shashi: Kenetic- mechanistic study of periodate oxidation of p-chloroaniline, Asian J. Chem. 15 (3&4) (2003), 1485-1490.
12. R. D. Kaushik, A.K. Chaubey and P. K. Garg: Kinetics and mechanism of periodate oxidation of p-phenetidine, Asian J. Chem. 15 (3&4) (2003), 1655-1658.
13. R.D. Kaushik, Mukta Dubey and R. K. Arya : Kinetics and mechanism of oxidation of 3, 5-di methylaniline in acetone-water medium, J. Curr Sci. 3 (1) (2003), 119 - 124.
14. R. D. Kaushik, Amrita, Mukta Dubey and R. P. Singh: Periodate oxidation of p-bromoaniline in acetone-water medium-a kinetic-mechanistic study, Asian J. Chem. 16 (2004), In Press.
15. R. D. Kaushik, Shashi, Amrita and Sumitra Devi: Kinetics and mechanism of Mn (II) catalysed periodate oxidation of 4-chloro-2-methylaniline, Asian J. Chem. 16 (2004), In Press.
16. N. Nalwaya, A. Jain and B. L. Hiran: Kinetics of oxidation of glycine by pyridinium bromochromate in acetic acid medium, J. Indian Chem. Soc. 79 (7) (2002), 587-589.

17. V. K. Pavolva, Ya. S. Sevchenko and K.B. Yatsmiriskii: Kinetics and mechanism of oxidation of diethylaniline by periodate, *Zh. Fiz. Khim.*, 44 (3) (1970), 658-63.
18. J. P. Phillips and F. C. Nachod (Ed.): *Organic Electronic Spectral Data*, Interscience publishers. N. York, Vol. 4 (1958), 108.
19. R. M. Silverstein : *Spectrometric identification of organic compounds*, 5th ed., John Wiley and Sons. Inc. N.Y., (1991).
20. S. P. Srivastava, M.C. Jain and R.D. Kaushik : Kinetics of Periodate oxidation of aromatic amines-oxidation of N, N- dimethylaniline, *Nat. Acad. Sci. Letters*, 2 (2) (1979), 63-64.
21. S. P. Srivastava, V. K. Gupta, M.C. Jain, M.N. Ansari and R.D. Kaushik: Thermodynamic and LFER studies for the oxidation of anilines by periodate ion, *Thermo Chimica Acta*, 68(1983), 27-33.

प्राकृतिक एवं भौतिकीय विज्ञान शोध पत्रिका

JOURNAL OF NATURAL & PHYSICAL SCIENCES

खण्ड 18 अंक 2, 2004

Vol. 18 No. 2, 2004



गुरुकुल कांगड़ी विश्वविद्यालय, हरिद्वार
Gurukula Kangri Vishwavidyalaya, Haridwar

शोध पत्रिका पटल

अध्यक्ष	स्वतन्त्र कुमार कुलपति	President	Swatantra Kumar Vice Chancellor
उपाध्यक्ष	वीरेन्द्र अरोड़ा प्राचार्य	Vice President	Virendra Arora Principal
सचिव	ए.के. चौपड़ा कुलसचिव	Secretary	A.K. Chopra Registrar
सदस्यगण	एस.पी. सिंह वित्ताधिकारी	Members	S.P. Singh Finance Officer
	वीरेन्द्र अरोड़ा प्रधान सम्पादक		Virendra Arora Chief Editor
	जे. पी. विद्यालंकार व्यवसाय प्रबन्धक		J. P. Vidyalankar Business Manager
	पी.पी. पाठक प्रबंध संपादक		P. P. Pathak Managing Editor

परामर्शदाता मण्डल

एस.एल. सिंह (सेवा निवृत्त प्रोफेसर एवं प्राचार्य)
योल जे चो (ज्योगसांग नेशनल विंविं, कोरिया)
एस. एन. मिश्रा (यूनिटरा, द. अफ्रीका)
सोम नैम्पल्ली (इमिरिट्स प्रोफेसर, कनाडा)
एस. पी. सिंह (मेमोरियल विवि., कनाडा)
देवकी एन. तलवार (इण्डियाना विवि., यू.एस.ए.)

Advisory Board

S. L. Singh E-mail : vedicmri@sancharnet.in
Yeol Je Cho, E-mail : yjcho@nongae.gsnu.ac.kr
S. N. Mishra, E-mail : mishra@getafix.ultr.ac.za
Som Naipally, E-mail : sudha@accglobal.net
S. P. Singh, E-mail : spsingh@math.mun.ca
Devaki N. Talwar, E-mail : talwar@iup.edu

सम्पादक मण्डल

वीरेन्द्र अरोड़ा (गणित)
ए.के. चौपड़ा (जन्तु एवं पर्यावरण विज्ञान)
आर. डी. कौशिक (रसायन शास्त्र)
वी. कुमार (कम्प्यूटर विज्ञान)
पुरुषोत्तम कौशिक (वनस्पति एवं सूक्ष्म जीव विज्ञान)
पी. पी. पाठक (भौतिकी) प्रबन्ध सम्पादक
वीरेन्द्र अरोड़ा प्रधान सम्पादक

EDITORIAL BOARD

Virendra Arora (Mathematics)
A.K. Chopra (Zoology & Env. Science)
R. D. Kaushik (Chemistry)
V. Kumar (Computer Science)
Purushottam Kaushik (Botany & Micro.)
P. P. Pathak (Physics) Managing Editor
Virendra Arora Chief Editor

A NOTE ON GENERALIZATIONS OF MIDPOINT INEQUALITY

B.G. PACHPATTE*

(Received 29.01.2004)

Abstract

In this note we establish two new generalizations of the midpoint inequality involving two functions and their second order derivatives by using a fairly elementary analysis.

Mathematics Subject classification (1991) : 26D15, 41A55

Key words and phrases : Midpoint inequality, second order derivatives, integration by parts, Hölder's inequality.

INTRODUCTION

The following classical inequality is well known in the literature as the midpoint inequality:

$$\left| \int_a^b f(x)dx - (b-a)f\left(\frac{a+b}{2}\right) \right| \leq \frac{\|f''\|_{\infty}}{24} (b-a)^3, \quad (1)$$

where the mapping $f: [a, b] \subset R \rightarrow R$ is twice differentiable on the interval (a, b) and having the second derivative bounded on (a, b) , i.e. $\|f''\|_{\infty} = \sup_{x \in (a, b)} |f''(x)| < \infty$.

Recently, Dragomir, et al. [1] have established the following inequality containing an integral inequality (1).

Let $f: [a, b] \subset R \rightarrow R$ be twice differentiable mapping on (a, b) and $f'': (a, b) \rightarrow R$ is integrable on (a, b) . Then

$$\left| \int_a^b f(x)dx - (b-a)f\left(\frac{a+b}{2}\right) \right| \leq \begin{cases} \frac{(b-a)^3}{24} \|f''\|_{\infty}, & \text{if } f'' \in L_{\infty}(a,b), \\ \frac{(b-a)^{2+\frac{1}{p}}}{8(2p+1)^{\frac{1}{p}}} \|f''\|_q, & \text{if } f'' \in L_q(a,b) \\ \quad \text{where } \frac{1}{p} + \frac{1}{q} = 1, p > 1, \\ \frac{(b-a)^2}{8} \|f''\|_1, & \text{if } f'' \in L_1(a,b). \end{cases} \quad (2)$$

For other results related to the midpoint inequality, see Chapter XV of the book by Mitrinovic et al. [2] The main purpose of this note is to establish two new generalizations of the midpoint inequality involving two functions and their second order derivatives. The analysis used in the proofs is elementary and our results provide new estimates on inequalities of this type.

STATEMENT OF RESULTS

Our main results are established in the following theorems.

THEOREM 1. Let $f, g: [a,b] \subset R \rightarrow R$ be twice differentiable mappings on (a,b) . Suppose that $f'', g'': (a,b) \rightarrow R$ are integrable on (a,b) . Then

$$\begin{aligned} & \left| 2 \left(\int_a^b f(x)dx \right) \left(\int_a^b g(x)dx \right) - (b-a) \left[f\left(\frac{a+b}{2}\right) \int_a^b g(x)dx + g\left(\frac{a+b}{2}\right) \int_a^b f(x)dx \right] \right| \\ & \leq \begin{cases} \frac{(b-a)^3}{24} \left[\|f''\|_{\infty} \int_a^b |g(x)|dx + \|g''\|_{\infty} \int_a^b |f(x)|dx \right], & \text{if } f'', g'' \in L_{\infty}(a,b), \\ \frac{(b-a)^{2+\frac{1}{q}}}{8(2q+1)^{\frac{1}{q}}} \left[\|f''\|_p \int_a^b |g(x)|dx + \|g''\|_p \int_a^b |f(x)|dx \right], & \text{if } f'', g'' \in L_p(a,b), \\ \quad \text{where } \frac{1}{p} + \frac{1}{q} = 1, q > 1, \\ \frac{(b-a)^2}{8} \left[\|f''\|_1 \int_a^b |g(x)|dx + \|g''\|_1 \int_a^b |f(x)|dx \right], & \text{if } f'', g'' \in L_1(a,b). \end{cases} \quad (3) \end{aligned}$$

Theorem 2. Let f, g, f'', g'' be as in Theorem 1. Then

$$\left| g\left(\frac{a+b}{2}\right) \int_a^b f(x)dx + f\left(\frac{a+b}{2}\right) \int_a^b g(x)dx - 2(b-a)f\left(\frac{a+b}{2}\right)g\left(\frac{a+b}{2}\right) \right| \leq \begin{cases} \frac{(b-a)^3}{24} \left[\|f''\|_\infty \left| g\left(\frac{a+b}{2}\right) \right| + \|g''\|_\infty \left| f\left(\frac{a+b}{2}\right) \right| \right], & \text{if } f'', g'' \in L_\infty(a, b), \\ \frac{(b-a)^{2+\frac{1}{q}}}{8(2q+1)^{\frac{1}{q}}} \left[\|f''\|_p \left| g\left(\frac{a+b}{2}\right) \right| + \|g''\|_p \left| f\left(\frac{a+b}{2}\right) \right| \right], & \text{if } f'', g'' \in L_p(a, b), \\ \frac{(b-a)^2}{8} \left[\|f''\|_1 \left| g\left(\frac{a+b}{2}\right) \right| + \|g''\|_1 \left| f\left(\frac{a+b}{2}\right) \right| \right], & \text{if } f'', g'' \in GL_1(a, b), \end{cases} \quad (4)$$

where $\frac{1}{p} + \frac{1}{q} = 1, q > 1,$

PROOFS OF THEOREMS 1 AND 2

Form the hypotheses and using the integration by parts formula, the following identities hold (See [1, p.26]):

$$\int_a^b f(x)dx - (b-a)f\left(\frac{a+b}{2}\right) = \int_a^b k(x)f''(x)dx, \quad (5)$$

$$\int_a^b g(x)dx - (b-a)g\left(\frac{a+b}{2}\right) = \int_a^b k(x)g''(x)dx, \quad (6)$$

where

$$k(x) = \begin{cases} \frac{(x-a)^2}{2} & x \in \left[a, \frac{a+b}{2}\right] \\ \frac{(b-x)^2}{2} & x \in \left[\frac{a+b}{2}, b\right]. \end{cases} \quad (7)$$

Multiplying both sides of (5) and (6) by $\int_a^b g(x)dx$ and $\int_a^b f(x)dx$ respectively and adding the resulting identities we have

$$\begin{aligned}
 & 2 \left(\int_a^b f(x)dx \right) \left(\int_a^b g(x)dx \right) - (b-a) f\left(\frac{a+b}{2}\right) \int_a^b g(x)dx \\
 & \quad - (b-a) g\left(\frac{a+b}{2}\right) \int_a^b f(x)dx \\
 &= \left(\int_a^b g(x)dx \right) \left(\int_a^b k(x)f''(x)dx \right) + \left(\int_a^b f(x)dx \right) \left(\int_a^b k(x)g''(x)dx \right). \tag{8}
 \end{aligned}$$

From (8) and using the properties of modulus we have

$$\begin{aligned}
 & \left| 2 \left(\int_a^b f(x)dx \right) \left(\int_a^b g(x)dx \right) - (b-a) \left[f\left(\frac{a+b}{2}\right) \int_a^b g(x)dx + g\left(\frac{a+b}{2}\right) \int_a^b f(x)dx \right] \right| \\
 & \leq \left(\int_a^b |g(x)|dx \right) \left(\int_a^b |k(x)| |f''(x)| dx \right) + \left(\int_a^b |f(x)|dx \right) \left(\int_a^b |k(x)| |g''(x)| dx \right). \tag{9}
 \end{aligned}$$

We consider the following three cases.

(i) If $f'', g'' \in L_\infty(a, b)$, then

$$\begin{aligned}
 & \int_a^b |k(x)| |f''(x)| dx \leq \|f''\|_\infty \int_a^b |k(x)| dx \\
 &= \|f''\|_\infty \left[\int_a^{\frac{a+b}{2}} \frac{(x-a)^2}{2} dx + \int_{\frac{a+b}{2}}^b \frac{(b-x)^2}{2} dx \right] = \frac{(b-a)^3}{24} \|f''\|_\infty. \tag{10}
 \end{aligned}$$

Similarly,

$$\int_a^b |k(x)| |g''(x)| dx \leq \frac{(b-a)^3}{24} \|g''\|_\infty. \tag{11}$$

Using (10) and (11) in (9) we get the first inequality in (3).

(ii) If $f'', g'' \in L_p(a, b)$ where $\frac{1}{p} + \frac{1}{q} = 1, q > 1$, then

$$\int_a^b |k(x)| |f''(x)| dx \leq \|f''\|_p \left(\int_a^b |k(x)|^q dx \right)^{\frac{1}{q}}$$

$$= \|f''\|_p \left(\int_a^{\frac{a+b}{2}} \left(\frac{(x-a)^2}{2} \right)^q dx + \int_{\frac{a+b}{2}}^b \left(\frac{(b-x)^2}{2} \right)^q dx \right)^{\frac{1}{q}} = \|f''\|_p \frac{(b-a)^{\frac{2+1}{q}}}{8(2q+1)^{\frac{1}{q}}} \quad (12)$$

Similarly,

$$\int_a^b |k(x)| |g''(x)| dx \leq \|g''\|_p \frac{(b-a)^{\frac{2+1}{q}}}{8(2q+1)^{\frac{1}{q}}} \quad (13)$$

Using (12) and (13) in (9), the second inequality in (3) holds.

(iii) If $f'', g'' \in L_1(a, b)$, then

$$\int_a^b |k(x)| |f''(x)| dx \leq \max_{x \in (a, b)} |k(x)| \|f''\|_1 = \frac{(b-a)^2}{8} \|f''\|_1. \quad (14)$$

Similarly,

$$\int_a^b |k(x)| |g''(x)| dx \leq \frac{(b-a)^2}{8} \|g''\|_1. \quad (15)$$

Using (14) and (15) in (9), we get the last inequality in (3). The proof of Theorem 1 is complete. In order to prove Theorem 2, we multiply both sides of (5) and (6) by

$g\left(\frac{a+b}{2}\right)$ and $f\left(\frac{a+b}{2}\right)$ respectively and add the resulting identities to get

$$g\left(\frac{a+b}{2}\right) \int_a^b f(x) dx + f\left(\frac{a+b}{2}\right) \int_a^b g(x) dx - 2(b-a) f\left(\frac{a+b}{2}\right) g\left(\frac{a+b}{2}\right)$$

$$= g\left(\frac{a+b}{2}\right) \int_a^b k(x)f''(x)dx + f\left(\frac{a+b}{2}\right) \int_a^b k(x)g''(x)dx. \quad (16)$$

From (16) and using the properties of modulus we have

$$\begin{aligned} &= \left| g\left(\frac{a+b}{2}\right) \int_a^b f(x)dx + f\left(\frac{a+b}{2}\right) \int_a^b g(x)dx - 2(b-a)f\left(\frac{a+b}{2}\right)g\left(\frac{a+b}{2}\right) \right| \\ &\leq \left| g\left(\frac{a+b}{2}\right) \int_a^b |k(x)| |f''(x)| dx + \left| f\left(\frac{a+b}{2}\right) \int_a^b |k(x)| |g''(x)| dx \right| \right| \end{aligned} \quad (17)$$

The rest of the proof can be completed by following the last arguments as in the proof of Theorem 1 below the inequality (9).

In concluding, we note that, in the special case, If we take $g(x) = 1$ and hence $g''(x) = 0$ in Theorems 1 and 2, then we get the inequality (2) which in turn contains the well known midpoint inequality given in (1).

REFERENCES

1. S.S. Dragomir, P. Cerone and A. Sofo, Some remarks on the midpoint rule in numerical integration, RGMIA research Report Collection, Vol.1, No. 2, 1998, pp 25-34.
2. D.S. Mitrinović, J.E. Pećarić and A.M. Fink, Inequalities for functions and their integrals and derivatives, Kluwer Academic Publishers, Dordrecht, 1994.

COMMON FIXED POINT THEOREMS FOR FOUR SELF-MAPS ON METRIC SPACES BY ALTERING DISTANCES

K.P.R. RAO* AND N. SRINIVASA RAO**

(Received 10.04.2004)

ABSTRACT

We obtain some common fixed point theorems for four self maps on metric spaces by alteration of distances using various definitions and conditions.

Key words: Fixed point, coincidence point, partially commuting, f-continuous, f-compatible, reciprocally continuous, altering distances.

2000 Mathematics subject classification : 47H10, 54H25.

INTRODUCTION

We prove some fixed point theorems for four self-maps by altering distances using various definitions and conditions.

First we give some well-known definitions:

Through out this paper, unless otherwise stated, $R(f)$ is the range of f .

DEFINITION 1. (Naidu [6]) The pair (f, g) is said to be f -continuous at z if $\{ffx_n\}$ and $\{fgx_n\}$ converge to fz whenever $\{x_n\}$ is a sequence in X such that $\{fx_n\}$ and $\{gx_n\}$ converge to z .

DEFINITION 2. (Jessy[2]) The ordered pair (f, g) is said to be weak compatible at z if $\{gffx_n\}$ converges to fz whenever $\{x_n\}$ is a sequence in X such that $\{fx_n\}$ and $\{gx_n\}$ converge to z , and $\{fgx_n\}$ and $\{ffx_n\}$ converge to fz .

DEFINITION 3. (G.Jungck [4]) The pair (f, g) is said to be partially commuting at z or weakly compatible at z if $fz = gz$ provided there exists $w \in X$ such that $fw = gw = z$.

DEFINITION 4. (Pathak and Khan [8]) The pair (f, g) is said to be f-compatible if $d(fgx_n, ggx_n) \rightarrow 0$ as $n \rightarrow \infty$ whenever $\{x_n\}$ is a sequence in X such that $fx_n \rightarrow z$ and $gx_n \rightarrow z$ as $n \rightarrow \infty$.

DEFINITION 5. (Jungck [3]) The pair (f, g) is said to be compatible (also called asymptotically commuting by Tiwari & Singh [15] see also [1] & [14]) if $d(fgx_n, gfx_n) \rightarrow 0$ as $n \rightarrow \infty$ whenever $\{x_n\}$ is a sequence in X such that $fx_n \rightarrow z$ and $gx_n \rightarrow z$ as $n \rightarrow \infty$.

*Dept. of Applied Mathematics, Acharya N.U.P.G. Centre, Nuzvid-521201, A.P.

**Dept. of Mathematics, Narasaraopeta Engineering college, Narsaraopet 522601, A.P.

CC-0. In Public Domain, Gurukul Kangri Collection, Haridwar

$gx_n \rightarrow z$ as $n \rightarrow \infty$.

DEFINITION 6. (Pant [7]) The pair (f, g) is said to be reciprocally continuous on X if $\{fgx_n\}$ converges to fz and $\{gfgx_n\}$ converge to gz whenever $\{x_n\}$ is a sequence in X such that $fx_n \rightarrow z$ and $gx_n \rightarrow z$ as $n \rightarrow \infty$.

DEFINITION 7. (Naidu[6]) The ordered pair (f, g) is said to be weakly f -compatible at z if either $\{gfgx_n\}$ or $\{ggx_n\}$ converges to fz whenever $\{x_n\}$ is a sequence in X such that $\{fx_n\}$ and $\{gx_n\}$ converge to z , and $\{fgx_n\}$ and $\{ffx_n\}$ converge to fz .

Naidu [6] observed the following

REMARK 8.

1. If f is continuous at z , then (f, g) is f -continuous at z .
2. If (f, g) is compatible at z , then (f, g) is weak compatible at z .
3. If (f, g) is either weak compatible or f -compatible or g -compatible, then (f, g) is weakly f -compatible at z for any $z \in X$.
4. If (f, g) is weakly f -compatible at z , then it is partially commuting at z .

MAIN RESULTS

THEOREM 9. Let (X, d) be a metric space and $f, g, S, T: X \rightarrow X$ be such that

$$\begin{aligned} \psi(d(fx, gy)) &\leq \varphi(\max\{\psi(d(Sx, Ty)) + \psi(d(fx, Sx) + \psi(d(gy, Ty)), \\ &\quad \psi(d(fx, Sx)) + \psi(d(Sx, gy)), \psi(d(gy, Ty)) + \psi(d(fx, Ty))\}) \end{aligned} \quad (9.1)$$

for all $x, y \in X$, where $\varphi: \mathbb{R}^+ \rightarrow \mathbb{R}^+$, φ is monotonically increasing on \mathbb{R}^+ , $\varphi(t) < t$ for all $t > 0$, ψ is continuous and monotonically increasing on \mathbb{R}^+ and $\psi(t) = 0 \Leftrightarrow t = 0$.

Suppose that there is a sequence $\{x_n\}$ in X such that

$$fx_{2n} = Tx_{2n+1} (= y_{2n}, \text{ say})$$

$$gx_{2n+1} = Sx_{2n+2} (= y_{2n+1}, \text{ say}) \text{ for } n = 0, 1, 2, \dots$$

and $d(y_n, y_{n+1}) \rightarrow 0$ as $n \rightarrow \infty$.

(9.2)

Then $\{y_n\}$ is a Cauchy sequence in X .

Further assume that $\{y_n\}$ converges to some $z \in X$.

Then the following statements are true.

1. If $u \in X$ and $Su = fu = u$, then $u = z$.

2. If $v \in X$ and $Tv = gv = v$, then $v = z$.
3. If $Sz = fz$, then $fz = Sz = z$.
4. If $Tz = gz$, then $gz = Tz = z$.
5. If $Sz = z$, then $fz = z$.
6. If $Tz = z$, then $gz = z$.
7. If (f, S) is f -continuous at z and weakly f -compatible at z , then $fz = z$.
8. If (g, T) is g -continuous at z and weakly g -compatible at z , then $gz = z$.
9. If $z \in R(S)$ and (f, S) is partially commuting at z , then $fz = Sz = z$.
10. If $z \in R(T)$ and (g, T) is partially commuting at z , then $gz = Tz = z$.
11. If (f, S) is S -continuous at z and weakly S -compatible at z , then $fz = Sz = z$.
12. If (g, T) is T -continuous at z and weakly T -compatible at z , then $gz = Tz = z$.
13. If (f, S) is compatible at z and reciprocally continuous at z , then $fz = Sz = z$.
14. If (g, T) is compatible at z and reciprocally continuous at z , then $gz = Tz = z$.
15. If $fz = Sz = Tz$, then $fz = Sz = Tz = gz = z$.
16. If $gz = Sz = Tz$, then $fz = Sz = Tz = gz = z$.
17. Maps f and S have a unique common fixed point, if one of the following statements is true.
 - (i) $fz = Sz$,
 - (ii) (f, S) is S -continuous at z and weakly S -compatible at z ,
 - (iii) $z \in R(S)$ and (f, S) is partially commuting at z ,
 - (iv) (f, S) is compatible at z and reciprocally continuous at z .
18. Maps g and T have a unique common fixed point, if one of the following statements is true.
 - (i) $gz = Tz$,
 - (ii) (g, T) is T -continuous at z and weakly T -compatible at z ,
 - (iii) $z \in R(T)$ and (g, T) is partially commuting at z ,
 - (iv) (g, T) is compatible at z and reciprocally continuous at z .
19. z is the unique common fixed point of f and S , as well as g and T , if at

least one of the statements (i), (ii), (iii) and (iv) of Statement 17 and at least one of those of Statement 18 is true.

20. z is the unique common fixed point of f and S , as well as g and T , if $Sz = Tz$ and at least one of the statements (i), (ii), (iii) and (iv) of statement 17 or at least one of those of Statement 18 is true.

PROOF. Suppose that $\{y_{2n}\}$ is not Cauchy.

Then, since $\{d(y_n, y_{n+1})\}$ converges to zero, there exists $\varepsilon > 0$ and strictly increasing sequences $\{m_k\}$, $\{n_k\}$ of positive integers such that $m_k \geq n_k \geq k$, $(y_{2n_k}, y_{2n_k}) \geq \varepsilon$ and

$d(y_{2m_k}, y_{2m_k-2}) < \varepsilon$ for all $k = 1, 2, 3, \dots$

we have $\varepsilon \leq d(y_{2n_k}, y_{2m_k}) \leq d(y_{2n_k}, y_{2m_k-2}) + d(y_{2m_k-2}, y_{2m_k-1}) + d(y_{2m_k-1}, y_{2m_k})$

Letting $k \rightarrow \infty$, we get

$$\lim_{k \rightarrow \infty} d(y_{2n_k}, y_{2m_k}) = \varepsilon \quad (9.3)$$

Also $\lim_{k \rightarrow \infty} d(y_{2n_k+1}, y_{2m_k-1}) = \varepsilon$ and $\lim_{k \rightarrow \infty} d(y_{2n_k}, y_{2m_k-1}) = \varepsilon$

(from (9.2), (9.3) and triangle inequality)

From (9.1) we have

$$\begin{aligned} \psi(d(fx_{2n_k}, gx_{2n_k+1})) &\leq \varphi(\max\{\psi(d(Sx_{2n_k}, Tx_{2n_k+1}))+\psi(d(fx_{2n_k}, Sx_{2n_k}))+\psi(d(gx_{2n_k+1}, Tx_{2n_k+1}), \\ &\quad \psi(d(fx_{2m_k}, Sx_{2m_k}))+\psi(Sx_{2m_k}, gx_{2n_k+1})), \\ &\quad \psi(d(gx_{2n_k+1}, Tx_{2n_k+1}))+\psi(d(fx_{2m_k}, Tx_{2n_k+1}))\}), \end{aligned}$$

That is

$$\begin{aligned} \psi(d(y_{2m_k}, y_{2n_k+1})) &\leq \varphi(\max\{\psi(d(y_{2m_k-1}, y_{2n_k}))+\psi(d(y_{2m_k}, y_{2m_k-1}))+ \\ &\quad \psi(d(y_{2n_k+1}, y_{2m_k})), \psi(d(y_{2m_k}, y_{2m_k-1}))+\psi(d(y_{2m_k-1}, y_{2n_k+1})), \\ &\quad \psi(d(y_{2n_k+1}, y_{2n_k}))+\psi(d(y_{2m_k}, y_{2n_k}))\}). \end{aligned}$$

Letting $k \rightarrow \infty$, we get

$$\psi(\varepsilon) \leq \varphi(\max\{\psi(\varepsilon) + \psi(0) + \psi(0), \psi(0) + \psi(\varepsilon), \psi(0) + \psi(\varepsilon)\})$$

$$= \varphi(\psi(\varepsilon)) \quad (\text{since } \psi(0) = 0).$$

Since $\phi(t+) < t$ for all $t > 0$, we must have $\psi(\varepsilon) = 0$.

Since $\psi(t) = 0 \Rightarrow t = 0$, we have $\varepsilon = 0$.

This is a contradiction. Hence $\{y_{2n}\}$ is a Cauchy sequence.

Since $\{d(y_n, y_{n+1})\}$ converges to zero, it follows that $\{y_n\}$ is a Cauchy.

Suppose now that $\{y_n\}$ converges to z for some $z \in X$. From (9.1), we have

$$\psi(d(fu, gx_{2n+1})) \leq \varphi(\max \{\psi(d(Su, Tx_{2n+1})) + \psi(d(fu, Su)) + \psi(d(gx_{2n+1}, Tx_{2n+1}))\},$$

$$\psi(d(fu, Su)) + \psi(d(Su, gx_{2n+1})),$$

$$\psi(d(gx_{2n+1}, Tx_{2n+1})) + \psi(d(fu, Tx_{2n+1})),$$

Letting $n \rightarrow \infty$, we get

$$\psi(d(u, z)) \leq \varphi(\max \{\psi(d(u, z)) + \psi(0) + \psi(0), \psi(0) + \psi(d(u, z)), \psi(0) + \psi(d(u, z))\})$$

(since $Su = fu = u$).

Hence $d(u, z) = 0$. Therefore $u = z$. This proves (1).

(2) Follows as in (1).

(3) Suppose that $Sz = fz$.

By taking $x = z$, $y = x_{2n+1}$ in (9.1) and letting $n \rightarrow \infty$ we get.

$$\psi(d(fz, z)) \leq \varphi(\max \{\psi(d(fz, z)) + \psi(0) + \psi(0), \psi(0) + \psi(d(fz, z)),$$

$$\psi(0) + \psi(d(fz, z))\}).$$

Hence $\psi(d(fz, z)) = 0$. Hence $d(fz, z) = 0$. Therefore $fz = z$.

(4) Follows as in (3).

(5) Suppose that $Sz = z$.

By taking $x = z$, $y = x_{2n+1}$ in (9.1) and letting $n \rightarrow \infty$, we get

$$\psi(d(fz, z)) \leq \varphi(\max \{\psi(0) + \psi(d(fz, z)) + \psi(0), \psi(d(fz, z)) + \psi(0),$$

$$\psi(0) + \psi(d(fz, z))\}).$$

Hence $\psi(d(fz, z)) = 0$. Hence $d(fz, z) = 0$. Therefore $fz = z$.

(6) Follows as in (5).

(7) Suppose that (f, S) is f -continuous at z and weakly f -compatible at z .

Since $fx_{2n} \rightarrow z$, $Sx_{2n} \rightarrow z$ and (f, S) is f -continuous at z , it follows that $ffx_{2n} \rightarrow fz$ and $Sfx_{2n} \rightarrow fz$. Since (f, S) is weakly f -compatible at z , it follows that either $Sfx_{2n} \rightarrow fz$ or $SSx_{2n} \rightarrow fz$.

Case (i) : Suppose that $Sfx_{2n} \rightarrow fz$. From (9.1)

$$\begin{aligned}\psi(d(ffx_{2n}, gx_{2n+1})) &\leq \varphi(\max \{\psi(d(Sfx_{2n}, Tx_{2n+1})), \psi(d(ffx_{2n}, Sfx_{2n})) + \\ &\quad \psi(d(gx_{2n+1}, Tx_{2n+1})), \psi(d(ffx_{2n}, Sfx_{2n})) + \psi(d(Sfx_{2n}, gx_{2n+1})), \\ &\quad \psi(d(gx_{2n+1}, Tx_{2n+1})) + \psi(d(ffx_{2n}, Tx_{2n+1}))\}),\end{aligned}$$

Letting $n \rightarrow \infty$, we get.

$$\begin{aligned}\psi(d(fz, z)) &\leq \varphi(\max \{\psi(d(fz, z)) + \psi(0) + \psi(0), \psi(0) + \psi(d(fz, z)), \\ &\quad \psi(0) + \psi(d(fz, z))\}).\end{aligned}$$

Hence $\psi(d(fz, z)) = 0$ and $d(fz, z) = 0$, i.e. $fz = z$.

Case (ii) : Suppose that $SSx_{2n} \rightarrow fz$.

Then by taking $x = Sx_{2n}$, $y = x_{2n+1}$ in (9.1) and letting $n \rightarrow \infty$, as in case (i) we get $fz = z$.

(8) Follows as in (7).

(9) Since $z \in R(S) \exists w \in X$ such that $Sw = z$.

Now by taking $x = w$, $y = x_{2n+1}$ in (9.1) and letting $n \rightarrow \infty$ we get

$$\begin{aligned}\psi(d(fw, z)) &\leq \varphi(\max \{\psi(0) + \psi(d(fw, z)) + \psi(0), \psi(d(fw, z)) + \psi(0), \\ &\quad \psi(0) + \psi(d(fw, z))\}).\end{aligned}$$

Hence $\psi(d(fw, z)) = 0$, and $d(fw, z) = 0$ i.e. $fw = z$.

Since (f, S) is the partially commuting at z and $Sw = fw = z$, we have $fz = Sz$.

The remaining part follows from Statement (3).

(10) Follows as in (9).

(11) Suppose that (f, S) is S -continuous at z and weakly S -compatible at z .

Since $fx_{2n} \rightarrow z$, $Sx_{2n} \rightarrow z$ and (f, S) is S -continuous at z , it follows that $Sfx_{2n} \rightarrow Sz$ and

$SS_{2n} \rightarrow Sz$. Since (f, S) is weakly S -compatible at z , it follows that either $fSx_{2n} \rightarrow Sz$ or $ffx_{2n} \rightarrow Sz$.

Case (i) : Suppose that $fSx_{2n} \rightarrow Sz$

From (9.1), we have

$$\begin{aligned}\psi(d(fSx_{2n}, gx_{2n+1})) &\leq \phi(\max\{\psi(d(SSx_{2n}, Tx_{2n+1})) + \psi(d(fSx_{2n}, SSx_{2n})) + \psi(d(gx_{2n+1}, Tx_{2n+1})), \\ &\quad \psi(d(fS_{2n}, SSx_{2n})) + \psi(d(SSx_{2n}, gx_{2n+1})), \psi(d(gSx_{2n+1}, Tx_{2n+1})) + \psi(d(fSx_{2n}, Tx_{2n+1}))\}).\end{aligned}$$

Letting $n \rightarrow \infty$, we get.

$$\begin{aligned}\psi(d(Sz, z)) &\leq \phi(\max\{\psi(d(Sz, z)) + \psi(0) + \psi(0), \psi(0) + \psi(d(Sz, z)), \psi(0) \\ &\quad + \psi(d(Sz, z))\}).\end{aligned}$$

Hence $\psi(d(Sz, z)) = 0$, and $d(Sz, z) = 0$, i.e. $Sz = z$.

Since $Sz = z$, $z \in R(S)$.

Since (f, S) is weakly S -compatible at z , it follows that (f, S) is partially commuting at z . Since $z \in R(S)$ and (f, S) is partially commuting at z , from Statement (9) it follows that $fz = Sz = z$.

Case (ii) Suppose that $ffx_{2n} \rightarrow Sz$.

Now by taking $x = f x_{2n}$, $y = x_{2n+1}$ in (9.1) and letting $n \rightarrow \infty$, we get $Sz = z$.

The rest of the proof follows as in Case (i).

(12) Follows as in (11).

(13) Since (f, S) is compatible, follows that $d(fSx_{2n}, Sfx_{2n}) \rightarrow 0$ as $n \rightarrow \infty$ whenever $fx_{2n} \rightarrow z$ and $Sx_{2n} \rightarrow z$ as $n \rightarrow \infty$. Since (f, S) is reciprocally continuous at z , follows that $fSx_{2n} \rightarrow fz$ and $Sfx_{2n} \rightarrow Sz$. Hence from the compatibility and reciprocal continuity of (f, S) at z , we have $fz = Sz$.

The remaining follows from statement (3).

(14) Follows as in (13).

(15) Follows from statements (3) and (6).

(16) Follows from statements (4) and (5).

Statements (17), (18), (19) and (20) are trivial from the above proofs.

Now we state the following theorem, which is a corollary of Theorem 9.

THEOREM 10. Let (X, d) be a metric space and $f, g, S, T: X \rightarrow X$ be such that

$$\psi(d(fx, gy)) \leq \phi(\max \{\psi(d(Sx, Ty)), \psi(fx, Sx)), \psi(d(gy, Ty)),$$

$$\psi(d(fx, Ty)), \psi(d(Sx, gy))\}) \quad (10.1)$$

for all $x, y \in X$, where $\phi, \psi: \mathbb{R}^+ \rightarrow \mathbb{R}^+$, ϕ is monotonically increasing on \mathbb{R}^+ , $\phi(t+) < t$ for all $t > 0$, ψ is continuous and monotonically increasing on \mathbb{R}^+ and $\psi(t) = 0 \Rightarrow t = 0$.

Suppose that there is a sequence $\{x_n\}$ in X such that

$$fx_{2n} = Tx_{2n+1} \quad (= y_{2n}, \text{ say})$$

$$gx_{2n+1} = Sx_{2n+2} \quad (= y_{2n+1}, \text{ say}) \text{ for } n = 0, 1, 2, \dots$$

and $d(y_n, y_{n+1}) \rightarrow 0$ as $n \rightarrow \infty$.

Then $\{y_n\}$ is a Cauchy sequence in X .

Further assume that $\{y_n\}$ converges to some $z \in X$.

Then statements (1) to (20) of Theorem 9 are true here also.

REMARK 11. Theorem 10 is a generalization of Theorem 1 of Naidu & Rajendra Prasad [5], Theorem 2.1, 2.4 of Sastry et al. [13], Theorem 1 of Sastry et al. [10], Theorem 7 and 12 of Sastry et al. [11]. Also Theorem 2 of Rathore et al. [9] and Theorem 3.1 of Sastry et al. [12] are corollaries of Theorem 10.

Finally we state a theorem which is a slight variant of Theorem 9.

THEOREM 12. Let (X, d) be a metric space and $f, g, S, T: X \rightarrow X$ such that

$$\psi(d(fx, gy)) \leq \phi(\max \{\psi(d(Sx, Ty)), \psi(d(fx, Ty)), \psi(d(Sx, gy))\} \cdot$$

$$+ \psi(d(fx, Sx)) + \psi(d(gy, Ty)))$$

for all $x, y \in X$, where $\phi, \psi: \mathbb{R}^+ \rightarrow \mathbb{R}^+$, ϕ is monotonically increasing on \mathbb{R}^+ , $\phi(t+) < t$ for all $t > 0$, ψ is continuous and monotonically increasing on \mathbb{R}^+ and $\psi(t) = 0 \Rightarrow t = 0$.

Suppose that there is a sequence $\{x_n\}$ in X such that

$$fx_{2n} = Tx_{2n+1} \quad (= y_{2n}, \text{ say})$$

$$gx_{2n+1} = Sx_{2n+2} \quad (= y_{2n+1}, \text{ say}) \quad \text{for } n = 0, 1, 2, \dots$$

and $d(y_n, y_{n+1}) \rightarrow 0$ as $n \rightarrow \infty$.

Then $\{y_n\}$ is a Cauchy sequence in X .

Further, assume that $\{y_n\}$ converges to some $z \in X$.

Then the following statements are true.

1. If $u \in X$ and $Su = fu = u$ then $u = z$.
2. If $v \in X$ and $Tv = gv = v$, then $v = z$.
3. If $Sz = fz$, then $fz = Sz = z$.
4. If $Tz = gz$, then $gz = Tz = z$.
5. If (f, S) is f -continuous at z and weakly f -compatible at z , then $fz = z$.
6. If (g, T) is g -continuous at z and weakly g -compatible at z , then $gz = z$.
7. If (f, S) is S -continuous at z and weakly S -compatible at z , then $Sz = z$.
8. If (g, T) is T -continuous at z and weakly T -compatible at z , then $Tz = z$.
9. If (f, S) is compatible at z and reciprocally continuous at z , then $fz = Sz = z$.
10. If (g, T) is compatible at z and reciprocally continuous at z , then $gz = Tz = z$.
11. Maps f and S have unique common fixed point if one of the following statements is true.
 - (i) $fz = Sz$
 - (ii) (f, S) is compatible at z and reciprocally continuous at z .
12. Maps g and T have unique common fixed point if one of the following statements is true.
 - (i) $gz = Tz$
 - (ii) (g, T) is compatible at z and reciprocally continuous at z .
13. Z is unique common fixed point of f and S , as well as g and T , if at least one of the following statement (i) and (ii) of statement (11) and at least one of those of statement (12) are true.

ACKNOWLEDGEMENTS: The authors express their heartfelt thanks to Prof. S.V.R. Naidu for his valuable suggestions in the preparation of this paper.

REFERENCES

1. U.C. Gairola, S.L. Singh and S.H.M. Whitfield : fixd point theorems on product of compact metric space, Cemoust. Math. 28 (1995) 541-548.
2. A. Jessy : Studies in fixed points and quasi-guages, Ph.D, Thesis, I.I.T. Madras, 1991.
3. G. Jungck : Compatible mappings and common fixed points, Internat. J. Math. & Math. Sci. 9 (4) (1986) 771-779.
4. G. Jungck : Common fixed points for non-continuous non-self maps on nonmetric spaces, Far East J. Math Sci 4 (2) (1996) 199-215.
5. S.V.R. Naidu and J. Rajendraprasad : Common fixed points for four self-maps on a metric space, Indian J. Pure Appl. Math. 16 (10) (1985) 1089-1103.
6. S.V.R. Naidu : Fixed point theorems for four self-maps on a metric space by altering distance, J. Indian Math. Soc. (to appear).
7. R.P. Pant : A common fixed point theorem under a new condition, Indian J. Pure Appl. Math. 30 (2) (1999) 147-152.
8. H.K. Pathak and M.S. Khan : A comparison of various types of compatible maps and common fixed points, Indian J. Pure Appl. Math. 28 (4) (1997) 477-485.
9. M.S. Rathore, Mamata Singh, Satira Rathore and Naval Singh : Concepts of the set E_α and common fixed points, Bull. Cal. Math. Soc. 94 (2002) 259-270.
10. K.P.R. Sastry, S.V.R. Naidu, I.H.N. Rao and K.P.R. Rao : Common fixed points for asymptotically regular mapping, Indian. J. Pure Appl. Math. 15 (8) (1984) 849-854.
11. K.P.R. Sastry, S.V.R. Naidu, I.H.N. Rao and K.P.R. Rao : Common fixed points under asymptotic regularity condition, Nep. Math. Sc. Rep. 9 (12) (1984) 63-71.
12. K.P.R. Sastry, G.V.R. Babu and D. Narayana Rao : Fixed points theorems in complete metric space by using a continuous control function, Bull. Cal. Math. Soc. 91 (6) (1999) 493-502.
13. K.P.R. Sastry, S.V.R. Naidu, G.V.R. Babu and G.A. Naidu : Generalisation of common fixed point theorems for weakly commuting maps by altering distances, Tamkang. J. Math. 31 (3) (2000) 243-250.
14. S.L. Singh and S.N. Mishra : Remarks on recent fixed points theorems and applications to integral equations, Demost. Math. 34 (4) (2000) 847-857.
15. B.M.L. Tivari and S.L. Singh : a note on recent generalizations of Jungck contraction principle, J. UPGC Acad. Soc. 3 (1986) 13-18

HARMFUL BIOEFFECTS OF HIGH FREQUENCY ELECTROMAGNETIC FIELDS

VIJAI KUMAR*, R.P. VATS* AND P.P. PATHAK**

(Received 05.06.2004)

ABSTRACT

The influence of electromagnetic field (EMF) of different frequencies at different distances from the transmission tower is discussed. Human body is a dielectric and consists of complex matrices of electric and magnetic fields. When the body is exposed to high frequency radiation, the electric field is induced inside the body. It is found that the current density at the surface of skin and bone increase with frequency of radiation being exposed. High frequency radiation also produces the dangerous diseases as leukemia and cancer risk in childhood.

Key works: Current density, electromagnetic field, conductivity of skin and bone

INTRODUCTION

Due to excessive modernization human beings are always surrounded by electromagnetic environment. Broadcast and communication transmitters as well as home appliances like induction heaters, VDU and TV sets and medical equipments emit EMFs of different frequencies and intensities. Though, EMFs of low frequencies are not so injurious but all types of high frequency EMFs affect the human body in many ways causing many diseases such as leukemia, brain tumor, and cancer risk in childhood [5,6,12]. These radiations affect the tissues, cells and skin of human body by induction of the electric and magnetic fields inside the body on exposure to the radiation. It seems well suited to predict changes in thermo-regulatory responses that may result from deposition of EM energy into the body and has provided the basis for several such analysis [4, 14, 16]. The EM radiation influences thermoregulatory behaviour like other sources of heat Hardy and Guieu [7] has summarized its thermophysiological effects of human body. Similarly, work has been done on physical character of human body in relation to heat loss [3]. If only one part of body (such as head or trunk) is exposed to EM radiation, the change in selected ambient temperature range is governed by an integrated energy deposition over the whole body and not by energy deposited in some specific locus as the brain. Indeed the hypothalamus appears to play no more vital role in behavioural thermoregulation during EM exposure than what the other thermo sensitive regions of the body do [1]. Therefore interaction may be expected between the body and radiation resulting into many types of biological effects. Efforts have been made to describe this type of interaction by Polk [11] and Mason et al. [9]. In the present paper current density in

*Department of Physics, M.S. College, Saharanpur (U.P.)

**Department of Physics, Gurukul Kangri University, Haridwar
Guru Nanak Dev Mohan Gurukul Kangri Collection, Haridwar

human skin and bone due to electromagnetic environment have been studied.

MECHANISM

The EMF produces two type of effects that control the dielectric behaviour. One is the oscillation of free charges or ions and the other is the rotation of dipole molecules at the frequencies of the applied EM energy. The first gives rise to conduction currents with an associated energy loss due to electrical resistance of the medium with an associated dielectric loss due to viscosity. Here we study the current density in the skin and human body induced due to the radiation of 10 kW transmission towers. Propagation of these radiations in the atmosphere means the propagation of electric and magnetic fields around the tower. These fields decrease as inversely proportional to distance from the tower. Simultaneously the conductivity of skin varies with frequency of radiation. If σ is the conductivity and E is the exposed electric fields, the surface current density in the skin is given by

$$\mathbf{J} = \sigma \mathbf{E}$$

The magnitude of induced electric field at a distance r from the antenna tower emitting at 10 kW is given by Pathak et al. [10]

$$E = 774.6/r$$

RESULTS AND DISCUSSION

The conductivity of skin and bone are increased with radiation frequency of exposure to body reported by Johnson and Guy [8]. Using these data calculated values of current density induced on the skin of human body at distances 100 m, 300 m, 500 m and 1 km from the tower are shown in Table 1. Table 2 shows calculated values of current density induced on bones of human body at various distance of 100 m, 300 m, 500 m and 1 km from the tower. Current flow or current density may affect the flow of ions. Watchtel [15] observed shift in neuron firing rate response to injected transmembrane. He also injected low frequency current into the seawater surrounding the cell preparation through external electrodes and observed some variation in effects. Shwan [13] proposes that current density in the tissues is the most useful parameter to the hazards. But the exact mechanism is to be worked out as to how EMF produces the heat inside the biological material, which also breaks down the protective mechanisms of heat control inside the body [2].

The values of current densities on the skin and bone of human beings are also increased with frequency of radiation as shown in Tables 1 and 2, which affect the human body in various ways. Thus the presence of transmission tower is harmful for the public health and therefore should be installed far away from populated areas.

Table 1: Variation of conductivity and current density on the skin of human beings with high frequency radiation.

Frequency of radiation (MHz)	Conductivity (mho)	Current density at different distances from the tower			
		100m	300m	500m	1 Km.
10	0.625	0.0893	0.0298	0.0178	0.00893
27.12	0.662	0.1337	0.0445	0.0267	0.0133
40.68	0.693	0.1621	0.54	0.0323	0.1614
100	0.889	0.28	0.0934	0.056	0.028
200	1.28	0.5081	0.1694	0.1063	0.0506
300	1.37	0.5684	0.1894	0.1135	0.0567
433	1.43	0.6034	0.2031	0.12	0.06034
750	1.54	0.6622	0.02208	0.1349	0.662
915	1.6	0.7008	0.2337	0.1401	0.07008
1500	1.77	0.8053	0.2686	0.1612	0.805
2450	2.21	1.047	0.3491	0.2095	0.1045
3000	2.26	1.0938	0.3645	0.2187	0.1091
5000	3.92	1.979	0.6597	0.3959	0.1975
5900	4.73	2.421	0.8083	0.4848	0.2421
8000	7.65	4.2304	1.4106	0.846	0.4222
10000	10.3	5.7062	1.9034	1.1422	0.5706

Table 2 : Variation of conductivity and current density in the bones of human beings with high frequency radiation.

Frequency of radiation (MHz)	Conductivity (mho)	Current density at different distances from the tower			
		100m	300m	500m	1 Km.
27.12	27.05	28.564	9.5216	5.7129	2.853
40.68	32.07	45.747	15.238	9.1527	4.5714
100	47.5	116.802	38.931	23.355	11.67
200	60	175.38	58.458	35.37	17.52
300	69.3	209.078	69.708	41.787	20.893
433	77.95	238.293	79.446	47.658	23.501
750	93.9	287.052	95.702	57.41	28.31
915	101.3	309.674	103.244	61.934	30.541
1500	120.9	369.591	123.221	73.918	36.451
2450	154.7	479.26	159.774	95.821	47.895
3000	172	532.856	177.641	106.571	53.251
5000	235.5	729.579	243.224	145.915	72.91
5900	262	863.552	287.859	172.684	85.276
8000	343	1189.52	396.542	237.87	118.849
10000	436.5	1560.48	520.133	312.053	155.917

REFERENCES

1. E.R. Adair : Mirowave challenges to the themoregulatory system, in Electromagnetic Waves and Neurobehavioral Function, M.E.O'Connor and R.H. Lovely, Eds., Alan R. Liss, New York 1988,179.
2. E.R. Adair : Themoregulation in the presence of microwave fields, in Handbook of Biological Effects of Electromagnetic Fields, C. Polk and E. Postow , Eds., CRC Press Inc. Boca Raton, 1996.
3. I.G. Berglund : Characterizing the thermal environment is microwave and themoregulation, E.R. Adair, Ed, Academic Press, New York, 1983. 15.
4. A.F. Emery, R.E. Short, A.W. Guy, K.K. Kraning, and J.C. Lin : The numerical thermal simulation of the human body when undergoing exercise or nonionizing electromagnetic irradiation, Trans. Am. Soc. Mech. Eng. 284, 1976.
5. M. Feychting, and A. Ahlbom : Magnetic fields and cancer in people residing near Swedish high power lines. Am.J. Eypedmiol. 138 (1993) 467.
6. M. Feychting and A. Ahlbom : Magnetic fields and cancer in people residing near Swedish high power lines. Institutet for miljomedicin, Karolinska intitutet, 1992.
7. J.D. Hardy and J.D. Guieu : Interaction activity of preoptic units II. Hypothetical network, J. Physiol.(Paris) 63 (1991) 264.
8. C.C. Johnson, and A.W. Guy : Non-ionising electromagentic effect in biological materials and systems. Proc. IEEE, 18 (June 1972) 692-718.
9. P.A. Mason, W.D. Hurt, T.J. Walters. , J.A. D Andera, P. Gajsek, K.L. Ryam, K.I. Nelson and J.M. Zinax : IEEE Trans Microwave Theory and Tech. 48 (11) (2000) 2050-2058.
10. P.P. Pathak , V. Kumar, and R.P. Vats : Harmful electromagnetic environment near transmission tower, Indian J. Radio & Space Phys. 32 (2003) 238-241.
11. Polk C : Introduction, in Handbook of Biological Effects of Electromagnetic fields, C. Polk and E. Postow (Eds.) CRC press Inc. Boca Raton, 1996.
12. D.A. Savitz, E.M. John and R.C. Kelckner : Magnetic field exposure from electric appliances and childhood cancer, Am. J. Epidemiol. 131 (1990) 763.
13. H.P. Schwany : Interaction of microwave and radio frequency radiation with biological systems, IEEE Trans. Microwave Theory Tech, Special Issue on Biological Effects of Microwaves) MTT- 19 (1971) 146-152.
14. R.J. Spiegel, D.M. Deffenebuag, and J.E. Mann : Athermal model of the human body exposed to an electromagnetic field, Bioelectromagnetics, 1 (1980). 253.
15. H. Watchel : Firing pattern changes and transmembrane currents produced by low frequency fields in pacemaker neurons, in Proc 18th Annu Hansford Life Sci. Symp. Technical information center, U.S. Department of Energy, Richland, Washington 1978. 132.
16. W.L. Way, H. Kritkos, and H.P. Schwan : Thermoregulatory physiologic responses in the human body exposed to microwave radiation, Bioelectromagnetics, 2 (1981) 341.

METHODS FOR SOLVING CUBIC EQUATIONS

ANITA SHARMA* AND V.K. SHARMA*

(Received 06.05.2004 and in Revised form 17.07.2004)

ABSTRACT

In this paper different methods for solving cubic equation are given. The contribution of Indian and Foreign mathematicians in the development for solving cubic equation are also discussed. Specially Omar Khayyam's geometrical method for solving particular type of cubic equation and Vieta's method for solving a cubic equation by converting it into quadratic are also explained.

Key Words : Cubic, Quadratic, Real, Imaginary

Classification No. : 01A32

INTRODUCTION

An equation of the form $ax^3+bx^2+cx+d=0$ where a, b, c and d are real numbers and $a \neq 0$ is called real cubic equation.

The oldest cubic equation which is of the form $x^3 = k$ is due to Mehaxchumus (c. 350 B.C.) [7].

The next cubic equation is due to Archimedes (c. 200 B.C.) [7].

$$x^3 + c^2 b = c x^2$$

Eutocius (c. 560 A.D.) solved the cubic equation of Archimedes by finding the intersection of two conics [7].

$$x^2 = (a^2/c^2)y \text{ (parabola)}$$

$$y(c-x) = bc \text{ (hyperbola)}$$

Next development of cubic equation was due to Arabs and Persians. Almahani (c. 860 A.D.) solved the cubic equation of Archimedes but he contributed nothing new. In pursuing the solution of cubic equation, Omar Khayyam (c. 1100 A.D.) blended solid geometry and geometry of conics. He gave geometrical and graphical solutions [3, 4, 7, 8], which are as follows :

*Department of Mathematics and Statistics, Gurukul Kangri Vishwavidyalaya, Haridwar.
CC-0. In Public Domain. Gurukul Kangri Collection, Haridwar

(a) Geometrical solution

let $x^3 + a^2x = b$ — (1) be a cubic equation, which is the intersection of two curves

- (i) $x^2 = ay$ (Parabola)
- (ii) $x^2 + y^2 = (b/a^2) x$ (Circle)

Now there are four possibilities :

- (1) when a and b are positive particularly we illustrate $a=1$ and $b=2$

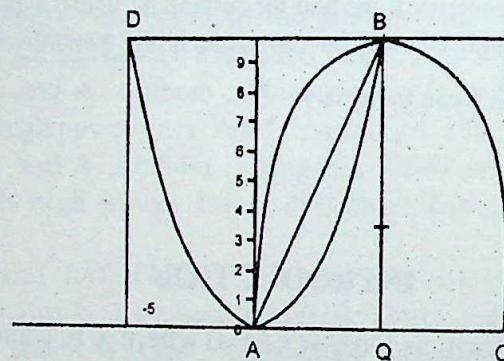


Fig. : 1

$AQ=1$ is the required solution of eqn (1).

- (2) $a = -1$ (negative) and $b = 2$ (positive), we must use the semicircle in quadrant 4

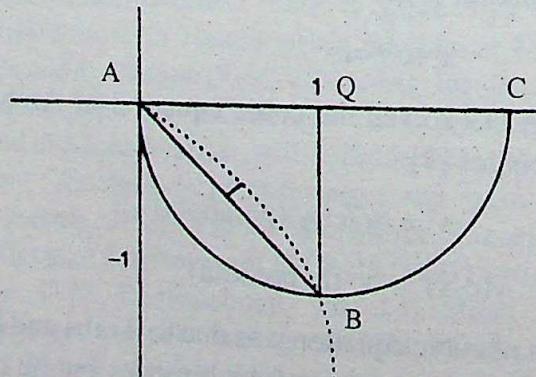


Fig. : 2

$AQ = 1$ is the required solution of eqn (1).

- (3) $a = -1$ (negative) and $b = -2$ (negative) because diameter b/a^2 can not be negative

so if b is negative, semicircle shifts to the left of the vertical axis.

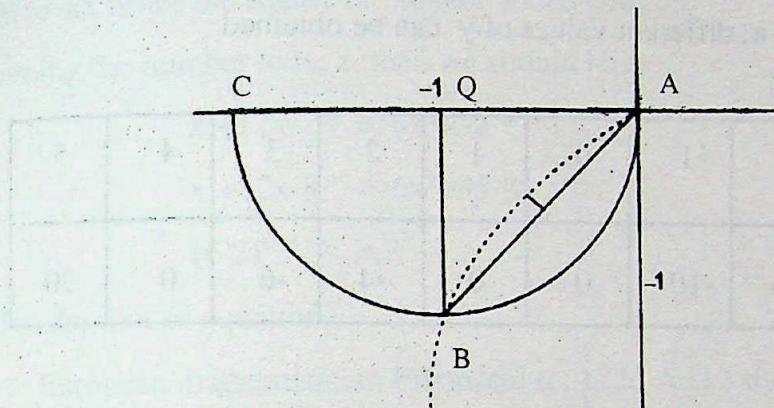


Fig. : 3

$AQ = -1$ is the solution of eqn(1).

(4) $a = 1$ (positive) and $b = -2$ (negative)

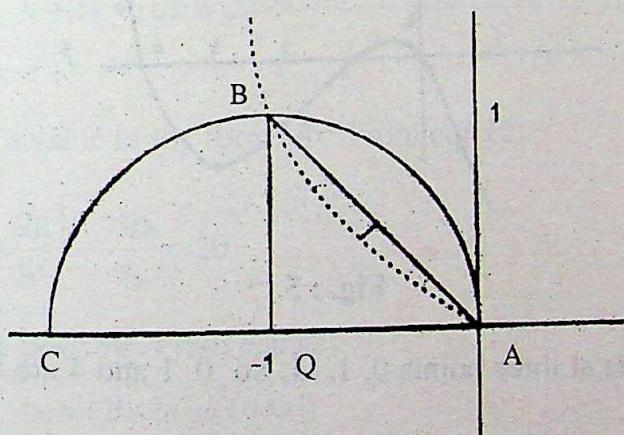


Fig. : 4

$AQ = -1$ is the solution of eqn(1).

(b) **Graphical solution**

Graphical solution of cubic equation is known from ancient time. It is also found in Omar Khayyam's work [1,4,6,10]. It is as follows :

Example: let $x^3 - 5x^2 + 4x = 0$ be a cubic equation

$$y = x^3 - 5x^2 + 4x$$

for different values of x , different values of y can be obtained

x	-2	-1	0	1	2	3	4	5
y	-36	-10	0	0	-4	-6	0	20

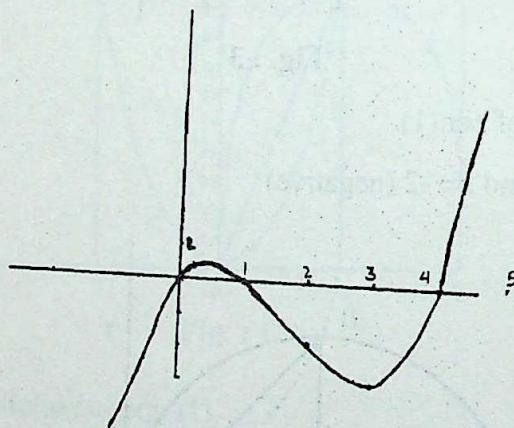


Fig. : 5

Graph cuts the x axis at three points 0, 1, 4. So 0, 1 and 4 are the roots of cubic equation.

Mahāvirāchārya (c 860 A.D.) considered equation of the following form by which only particular type of cubic could be solved [3,7]. It is as follows:

$$Ax^n = q$$

Bhāskarāchāryā (c. 1150 A.D.) wrote a problem of cubic equation in his Bijganita [3, 5, 7] which is as follows:

राशिद्विदिशनिषयो राशिधानाङ्क्यच्चकः समो यः स्यात्।

राशिकृतिः घट्गुणिताः पञ्चग्रिशाद्युता विद्वन् ॥ [5, P. 233]

This means

What is the number which multiplied by 12 and increased by the cube of number is equal to six times the square of number added with 35.

Assuming the number to be x , then we should have

$$x^3 + 12x = 6x^2 + 35$$

$$x^3 + 12x - 8 = 6x^2 + 35 - 8$$

$$(x-2)^3 = 3^3$$

$x = 5$ is the root of equation.

European mathematician Fibonacci (c. 1225 A.D.) studied cubic equation in his work *Felos* and *Liber Quadratorum* [2, 7, 9]. For the problem $x^3 + 2x^2 + 10x = 20$, he proved that x can not be irrational such as $\sqrt{a} + \sqrt{b}$.

Example: Consider the following cubic equation.

$$x^3 + 2x^2 + 10x = 20 \quad \text{---(2)}$$

since $2 < x < 1$

So, x must have a value of the form a/b . From eqn.(2)

$$\frac{a^3}{b^3} + \frac{2a^2}{b^2} + \frac{10a}{b} = 20$$

which can not be possible. From eqn.(2)

$$x = (20 - 2x^2) / (10 + x^2)$$

put $x = \sqrt{t}$

$$\sqrt{t} = (20 - 2t) / (10 + t)$$

which is impossible.

He also gave an approximate solution in sexagesimal form

$$x = 1, 22, 7, 42, 33, 4, 40$$

$$x = 1 + \frac{22}{60} + \frac{7}{60^2} + \frac{42}{60^3} + \frac{33}{60^4} + \frac{4}{60^5} + \frac{40}{60^6}$$

In decimal notation

$$x \sim 11368808106$$

Scipio del Ferro (1465 - 1526 A.D.) and Cardan published the general method for solving cubic equation by converting it into quadratic [1, 2, 9]. Let the cubic equation be

$$c_1 x^3 + c_2 x^2 + c_3 x + c_4 = 0 \quad \dots \dots \dots (3)$$

$$\text{or } x^3 + px^2 + qx + r = 0 \quad \dots \dots \dots (4)$$

$$\text{where } p = c_2/c_1, \quad q = c_3/c_1, \quad r = c_4/c_1$$

$$\text{put } x = y - p/3 \text{ in eqn. (4)}$$

$$\text{we get } y^3 + Ay + B = 0 \quad \dots \dots \dots (5)$$

$$\text{where } A = \frac{1}{3}(3q - p^2)$$

$$\text{and } B = \frac{1}{27}(2p^3 - 9pq + 27r)$$

$$y = m^{1/3} + n^{1/3} \text{ be the solution of eqn. (5)}$$

$$y^3 - 3m^{1/3}n^{1/3} - (m+n) = 0 \quad \dots \dots \dots (6)$$

comparing eqn. (5) and eqn. (6)

$$m^{1/3}n^{1/3} = -A/3$$

$$m + n = -B$$

If m and n are the roots of quadratic equation then equation is-

$$t^2 + Bt - A^3/27 = 0$$

$$t = \frac{-B \pm \sqrt{B^2 + 4A^3/27}}{2} \quad \dots \dots \dots (7)$$

taking cube root of m and n, we get value of y

$$y = \left[\frac{-B}{2} + \left(\frac{B^2}{4} + \frac{A^3}{27} \right)^{1/2} \right]^{1/3} - \left[\frac{B}{2} + \left(\frac{B^2}{4} + \frac{A^3}{27} \right)^{1/2} \right]^{1/3} \quad \dots \dots \dots (8)$$

three conditions are possible for the roots of equation

- 1) $B^2 + 4A^3/27 < 0$, all roots are real but eqn (8) gives them in an imaginary form, thus Cardan's solution is not convenient.
- (2) $B^2 + 4A^3/27 > 0$, one root is real, other two are pair of complex conjugate.
- (3) $B^2 + 4A^3/27 = 0$, cubic consists two equal roots.

Cardan also proved this formula geometrically. The proof of formula is given in the end of paper.

In 1545 A.D. Cardan published the solution of following equation [9]

$$x^3 + mx = n \quad \dots \dots \dots (9)$$

$$\text{and} \quad (a-b)^3 + 3ab(a-b) = a^3 - b^3 \quad \dots \dots \dots (10)$$

comparing eqn. (9) and eqn. (10), we get

$$3ab = m$$

$$a^3 - b^3 = n$$

then $(a-b)$ is one of the solution of eqn. (9).

Vieta (b. 1540 A.D.) and Harriat (b. 1560 A.D.) published following method for solving cubic equation (removing x^2 form) [2, 7]

$$x^3 + 3b^2x = 2c^3 \quad \dots \dots \dots (11)$$

put

$$x = (e^2 - b^2)/e \text{ in eqn. (11)}$$

we get

$$e^6 - 2c^3e^3 = b^6 \quad \dots \dots \dots (12)$$

this equation is quadratic in e^3 and then $x = e - b^2/e$, e , and b^2/e are the roots of expression (11).

Vieta (b. 1540 A.D.) and Girard (b. 1629 A.D.) compared the cubic eqn. (9)

with trigonometrical identity [4, 7]

$$\cos^3 \theta - \frac{3}{4} \cos \theta = \frac{1}{4} \cos 3\theta \quad \dots \dots \dots \quad (13)$$

Example: Find depth z to which a spherical solid ball made of cork (sp. qr. $1/4$) will sink into water

Solution: let radius of ball be a . By Archimedes principle

$$z^3 - 3az^2 + a^3 = 0 \quad \dots \dots \dots \quad (14)$$

put $z = nx + a$ in eqn. (14)

we get $n^3 x^3 - 3a^2 nx - a^3 = 0 \quad \dots \dots \dots \quad (15)$

for $x = \cos \theta$ eqn. (13) is identical with eqn. (15)

$$n = 2a \text{ and } \cos 3\theta = 1/2$$

for three distinct values of θ $20^\circ, 100^\circ, 140^\circ$ three values of z can be obtained

$$\text{Required depth } z = a + 2a \cos 100^\circ$$

This method is suitable when all roots are real, for imaginary roots identity involving hyperbolic function can be used.

Bomballi (c. 1572 A.D.) introduced complex numbers by which the solution of irreducible case could be found by Cardan rule. It was also given by Kastner (b. 1745 A.D.) and Chariout (b. 1746 A.D.) [2, 7]. Following is an example :

Example: let $x^3 - 63x - 162 = 0 \quad \dots \dots \dots \quad (16)$

be a cubic eqn. to solve

from eqn. (8)

$$x = [(81 + 30\sqrt{-3})]^{1/3} - [(81 - 30\sqrt{-3})]^{1/3}$$

taking cube root $x = -6$ is one of the root of given cubic equation. Dividing the eqn. (16) by $(x+6)$, we get

$$x^2 - 6x - 27 = 0$$

where roots are 9 and -3

It means three roots of cubic equation are -6, 9, -3.

In 1668, Hudde gave another method for solving cubic equation [7]

let $x^3 = qx + r$ ----- (17)

put $x = y+z$ in eqn.(17)

$$y^3 + z^3 + 3yzx = qx + r \text{ ----- (18)}$$

comparing coefficients of like terms

we get $y^3 + z^3 = r$

$$3yz = q$$

solving for y and z we get value of x.

Newton (c.1669 A.D.) suggested that a particular root of a equation could be found by substituting a known approximate root in the given equation [4, 7]. Following is an example :

Example:- let the cubic equation be

$$y^3 - 2y - 5 = 0 \text{ ----- (19)}$$

for $y = 2$

$$y^3 - 2y - 5 = -1$$

for $y = 3$

$$y^3 - 2y - 5 = 16$$

$$2 < y < 3$$

let $y = 2 + p$ ----- (20)

from eqn. (19)

$$(2 + p)^3 - 2p(2+p) - 5 = 0$$

$$p^3 + 6p^2 + 10p - 1 = 0 \text{ ----- (21)}$$

we get $p = 0.1$ approximately

put $p = 0.1 + q$ ----- (22)

we have

$$q^3 + 6.3q^2 + 11.23q + 0.061 = 0 \quad (23)$$

$$q = -0.0054 \text{ approximately}$$

put $q = -0.0054 + r$ ----- (24)

from eqn. (23) (omitting r^3 term)

$$6.3r^2 + 11.1619r + .000541708 = 0 \quad (25)$$

$$r = -0.00004854 \text{ approximately}$$

$$r = -0.00004854 + s \quad (26)$$

from eqn. (25) (omitting s^2 term because s^2 is very small real number)

$$177076885 s = 1.2584$$

$$s = 0.00000000 7106 \text{ approximately}$$

$$s = 0.00000000 7106 + t$$

being a very small real number we will not consider the value of t

Put the value of s in eqn. (26)

we get

$$r = -0.000048532894$$

from eqn. (24)

$$q = -0.005448532894$$

from eqn. (22)

$$p = 0.09451467106$$

and finally we have

$$y = 2.09451467106 \text{ (approximately)}$$

The process terminates if the root is commensurable and it may be carried to any required number of decimal places if it is incommensurable.

De Moivre's (1667-1754 A.D.) method for solving cubic equation was based on the identity [1,4]

$$(\cos \theta + i \sin \theta)^n = \cos n\theta + i \sin n\theta$$

let the cubic equation be $ax^3 + bx^2 + cx + d = 0$ and can be written as

$$(x + b/3a)^3 = (b^3/27) a^3 - d/a \quad \dots \dots \dots (27)$$

$$\text{if } b^2 = 3ac$$

Example : let the cubic equation be

$$z^3 + 6z^2 - 12z - 12 = 0, \text{ from eqn. (27) we get}$$

$$(z-2)^3 = 4$$

$$(z-2)^3 = 4(\cos 2r\pi + \sin 2r\pi)$$

$$(z-2) = 4^{1/3}(\cos 2r\pi + \sin 2r\pi)^{1/3}$$

for $r = 0, 1, 2$ we get three roots. De Moivre's method can be applied when R.H.S. of eqn. (22) is not real. Pure algebraic method fails to find the cube root of general complex number in explicit form. When $b^2 = 3ac$ is not satisfied, this method can be used only particular cases.

Lagrange (1736-1813 A.D.) introduced a method of expressing the root of numerical equation in the form of continued fraction [1].

Example: $x^3 - 2x - 5 = 0 \quad \dots \dots \dots (28)$

root lie between 2 and 3

put $x = 2 + 1/y$ in eqn. (28)

transformed eqn. is

$$y^3 - 10y^2 - 6y - 1 = 0 \quad \dots \dots \dots (29)$$

root lie between 10 and 11

put $y = 10 + 1/z$ in eqn. (29), we get

$$61z^3 - 94z^2 - 20z - 1 = 0 \quad \dots \dots \dots (30)$$

$$1 < z < 2$$

put

$$z = 1 + 1/u \text{ in eqn. (30)}$$

$$54u^3 + 25u^2 - 89u - 61 = 0$$

$$1 < u < 2 \text{ and so on-----}$$

expression for the root

$$x = 2 + \cfrac{1}{10 + \cfrac{1}{1 + \cfrac{1}{1 + \cfrac{1}{\dots}}}}$$

Geometrical proof of formula

Cardan's geometrical proof of his formula given by eqn. (8) is as follows :

Let the cubic equation be $x^3 + mx = n$ ----- (31)

where m and n are positive because diamensions of cube be positive. Assuming a cube Z with side t . Where $t = (t-u) + u$

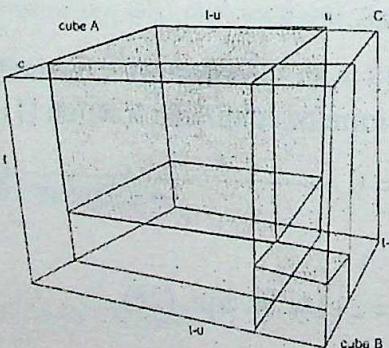


Fig. : 6

This cube of side t contains two cube A and B having sides $(t-u)$ and u and four cuboids.

vol. of cube Z = vol. of cube A + vol of cube B + vol. of four cuboids

$$t^3 = (t-u)^3 + u^3 + 2tu(t-u) + (t-u)u^2 + u(t-u)^2$$

$$t^3 - u^3 = (t-u)^3 + (t-u) 3tu \quad \dots \dots \dots (32)$$

let $x = t-u$ then

$$t^3 - u^3 = x^3 + 3tux \quad \dots \dots \dots (33)$$

comparing eqn. (31) and eqn. (33)

$$m = 3tu \quad \dots \dots \dots (34)$$

$$n = t^3 - u^3 \quad \dots \dots \dots (35)$$

from eqn. (34) and eqn. (35), we get

$$\text{it means } t^6 - nt^3 - m^3/27 = 0 \quad \dots \dots \dots (36)$$

$$t = [+n/2 + (n^2/4 + m^3/27)^{1/2}]^{1/3}$$

from eqn. (35)

$$u = [-n/2 + (n^2/4 + m^3/27)^{1/2}]^{1/3}$$

and finally $x = t-u$ is the solution.

CONCLUSION

Omar Khayyam's, Bhāskarācharya's, and Eutocius's methods are suitable only for solving particular type of equations. Methods given by Ferro, Cardan, Bomballi, Vieta, Girard and Newton are general methods. Numerical methods based on convergence of sequences are now used for solving cubics.

REFERENCES

1. W. S. Burnside, A.W. Panton, Theory of Equations, S. Chand and Co. Delhi 1957.
2. J.N. Crossley, The Emergence of Number, World Scientific Publishing Co. Pte. Ltd. Singapore 1987.
3. B. Dutta and A.N. Singh, History of Hindu Mathematics, Asia Publishing House, Bombay, 1962.
4. R.C. Gupta, Many Methods of Solving a Cubic JOBIC, 1961-62, 54-63.
5. पं० देवचन्द्र ज्ञा "बीजगणितम्" कृष्णदास अकादमी चौखम्बा प्रेस वाराणसी। १६८३
6. Lancelot Hogben, Mathematics for the Million, George Allen and Unwin Ltd. Museum Street, London 1957.

7. D.E. Smith, History of Mathematics Vol. II, Dover Publications New York 1959.
8. G.S. Smirnova, Geometrical Solutions of Cubic Equation in Raffaele Bomballii's Algebra Istor, Metodol Estestv, Nauk 36, 1989, 123-129.
9. V.S. Vardhrajan, Algebra in Ancient and Modern Times, Hindustan book Agency, New Delhi 1997.
10. P.D. Yardley, Graphical Solution of Cubic Equation developed from the work of Omar Khayyam Bull Inst Math Appl 26 (5-6) 1990 122-125.

CONSTRUCTION OF SOME IMPORTANT VEDIS FROM SULBA SŪTRA

YOGITA BANA* AND V.K. SHARMA*

(Received 06.05.2004, Revised 19-07-2004.

ABSTRACT

Construction of some vedis from Sulba Sūtra are given.

Key words and phrases: Prāci, Sacrifice, Oblong, Citi, Isosceles trapenziun, Altar, Kāmya Agni, Nitya Agni,

Classification Number: 01A32

INTRODUCTION

The science of Mathematics which is said to be the " queen of all sciences" or " the peak of all sciences" has originated in so hoary manner that it is impossible to trace its origin in any of the written records of the human civilization. Mathematics was known in ancient India as Ganir-veda or the Veda (i.e. Knowledge) of calculations for detail description [15].

Only seven Sulba Sūtra are available at present. These are : 1 Baudhāyana, 2. Āpastamba, 3. Vādhula, 4. Maitrāyana, 5. Manava, 6. Vārāha, 7. Kātyāyana. First six belong to Traitriya Samhitā of Kṛsna Yajurveda. Only the sūtra of Kātyāyana belongs to Śukla Yajur-veda. Out of these Baudhāyana, Āpastamba and Kātyāyana deal with the mathematical portion-namely Geometry, in detail Manava gives a very scanty description of the geometry. It deals only with the Square and Circle. The Baudhāyana, Āpastamba and Manava make use of these rules in the construction of various citis. But description gives only their dimensions accordingly to their shapes.

The most important points observed in all of them are:-

1. The shape of the citis or the altars of the sacred fires may differ but their areas will remain the same.
2. In the construction of citis, there will be five layers and each layer will consist of 200 bricks.
3. In the case of the three sacred fires only Garhapatya will have five layers and each layer will consist of 21 bricks [12].

* Department of Mathematics & Statistics, Gurukul Kangri Vishvavidyalaya, Haridwar.
CC-0. In Public Domain. Gurukul Kangri Collection, Haridwar

GLIMPSES OF FIRE-ALTARS

There is fundamental difference between the two types of ritual. The grahya ritual is performed by the householder and his wife in the Gārhapatya fire only. In this ritual, as mentioned in Rv. X 8.5.27 "In this thy (husband's family) affection may increase with offspring, be watchful over the Garhapatya Agni in this house." A priest is seldom required. The srauta ritual is performed by the sacrificer and his wife in more than one fire viz. Gārhapatya, Āhavaniya and Daksina with important role of priest. However there are various views in concern of different agnis on the basis of seasonal and periodical sacrifices as well as Nitya and Kamya.

Kāmya agnis were having five layers, each made of fixed number of bricks. The Nitya fire has 21 brick and kamya fire has 200 bricks in each layer [10].

Name	Horizontal Section/Shape	Sūtra reference
Nitya Agni		
Āhavaniya	Square	Baudhāyana Sulba sūtra (BSS 7.4-7.7)
Gārhaptya	Circle	BSS 7.8
Daksinagni	Semi Circle	[13]

Kamya Agni

Optional fire altars which are built for the fulfilment of desire i.e. Kamyagni Taittariya-Samhita 5.4.11.1-3. These are:- (Each optional fire has an area $7\frac{1}{2}$ square purusa) (see 13).

Name	Horizontal Section/Shape	Sutra reference
Kamya Agni		
Caturasrasayenacit	Hawk bird with square body, wings and tail rectangles	BSS 8.1-8.18
Vakrapksavyasta-puccahasyenacit	Hawk bird with bent wings and out spread tail	9.2-9.10 BSS 10.1-10.14 11.2-11.13

Kankacit	Hawk bird with curved wing and tail	BSS 12.1-12.8
Alajacit	Alaja bird with curved wings and tail	BSS 13.1-13.5
Praugacit	Isosceles triangle	BSS 14.4-14.8
Ubhayata praugacit	Rhombus	BSS 15.2-15.6
Rathacakracit	Chariotwheel	BSS 16.3-16.5, 16.6-16.20
Dronacit	Square trough	BSS 17.1-117.12
	Circular trough	BSS 18.1-18.15
Samasanacit	Isosceles trapezium	BSS 19.1-19.11
Kurmacit	Tortoise	BSS 20.1-20.21 21.2-21.13
Vedi		
Māhavedi/	Isocels trapezium	BSS 4.3
Saumikyāvedi	(face $a =$, base $b = 30$, height $c = 36$)	[2] & ([3], p.69) ([7], p.22)
Sautramanivedi	Isosceles trapezium $(a = 8\sqrt{3}, b = 10\sqrt{3}, c = 12\sqrt{3})$	BSS 3.12, 1.85
Paitrkivedi	Isoceles trapezium $(a = 8, b = 10, c = 12)$	BSS 1.82, 3.11 ([17], p. 68)
Ultaravedi	Square having four corners in four corrdinate directions. Four cornered and each side is 36 angulas	
Asvamedhavedi	Four cornered and each side is 36 angulas Isosceles trapezium $(a = 32, b = 40, c = 54)$	BSS 3.10

(Units are in padas otherwise stated)

Following are some of the geometrical methods which are used for construction of altars.

1. Construction of a trapezium Similar to a given trapezium.
2. Construction of trapezium similar to double of its area.
3. Construction of a trapezium similar to third part of its area.
4. Construction of an isosceles trapezium similar to given trapezium by increasing/diminishing its sides by a certain proportion.
5. Construction of an isosceles trapezium of given area.
6. Construction of a square of given area
7. A quadrilateral formed by the lines joining the middle parts of the sides of a square is a square whose area is half that of the original one.
8. A quadrilateral formed by the lines joining the middle points of the sides of the rectangle is a rhombus whose area is half that of the rectangle.

In addition of geometrical problems, Sulb Sutras also concern extensively with the use of instruments such as bamboo rod, geometrical compass, peg, rope and sanku for altar construction; relative distance between Āhavaniya, Dakṣināgni, Gāṛhapatya and Uttaravedi; relative position of various altars with regard to Māhavedi; enlargement of fire altars; a few indeterminate problems; a little of geometrical algebra; fractions; elementary concept of arithmetic series [13].

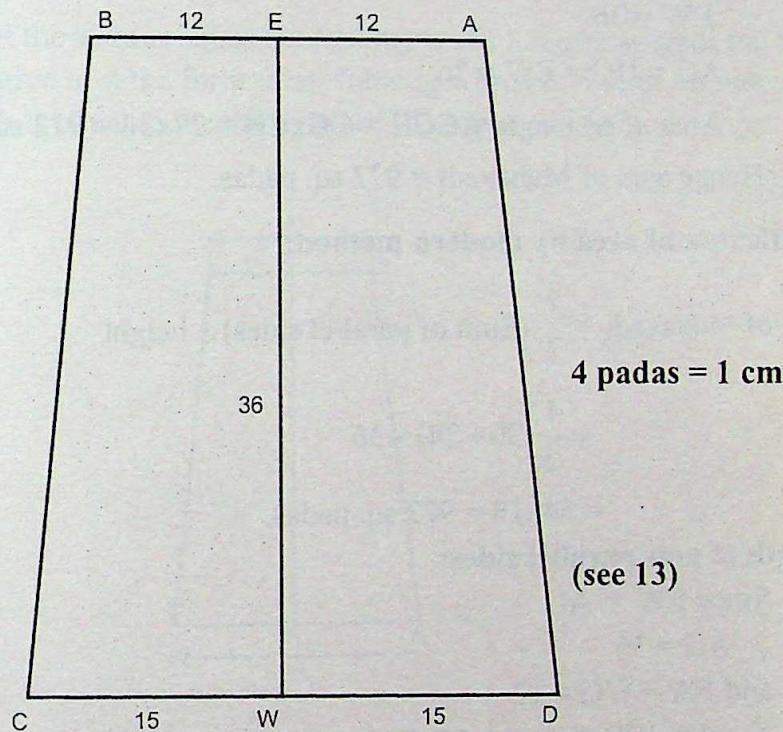
MAHĀVEDI

त्रिष्णशत्पदानि प्रक्रमा वा पश्चात्तिरश्ची भवति षट्त्रिष्णत्प्राची
चतुर्विष्णशतिः पुरस्तात्तिरश्चीति महावेदर्विज्ञयाते मानयोगस्तस्या व्याख्यातः॥ (BSS 1.90)

The meaning of above Sūtra is as follows:

"The western side is thirty padas of prakrams long, the prācī thirty-six the eastern side twenty four" this is the tradition for Mahāvedi (the Vedi used at the soma sacrifices); it has been explained how it is to be measured out [17].

In other words above may be stated as *Mahāvedi* described as isosceles trapezium measuring 80 padas or prakarmas on western side, 36 padas along east-west line 24 padas on eastern side [13].



Description :

Mahāvedi ABCD is isosceles trapezium of width 36 and parallel sides 24 and 30 units. This is constructed by first making the central-axis (east-west line), points W and E and then drawing the four segments WD, WC, EA and EB along the other cardinal directions (south and north) ([13], p.22) *Mahavedi* is called *Saumiki vedi* in ASS.

Calculation of area of *Mahavedi* by *Sulba Sūtra* method:

Draw a perpendicular from A on side CD and from C on AB. Let perpendiculars meet on WD at G and on AB at H.

Area of *Mahavedi* ABCD = Area of rectangle AGCH.

E and W are middle points of sides AB and CD.

$$\therefore BE = EA = 12$$

$$\therefore CW = WD = 15$$

$$\therefore WG = 12 \text{ and } CG = 15 + 12 = 27$$

$$\therefore EW = 36$$

$$\therefore AG = HC = EW = 36$$

$$\therefore \text{Area of rectangle } AGCH = CG \times CH = 27 \times 36 = 972 \text{ sq. padas [4].}$$

Hence area of Mahavedi = 972 sq. padas.

Verificaton of area by modern method:

$$\text{Area of Mahavedi} = \frac{1}{2} (\text{Sum of parallel sides}) \times \text{height}$$

$$= \frac{1}{2} (30 + 24) \times 36$$

$$= 54 \times 18 = 972 \text{ sq. padas.}$$

Length of non-parallel sides:

$$\text{Since } EW = AG$$

$$\therefore AG = 36$$

$$\text{and } EA = WG = 12$$

$$\therefore GD = WD - WG = 15 - 12 = 3$$

Now in triangle AGD, using Sulbakra's theorem

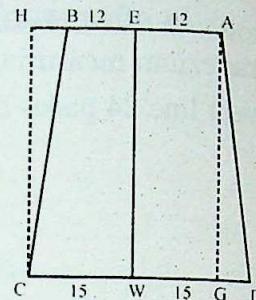
$$AD^2 = AG^2 + GD^2$$

$$\therefore AD^2 = (36)^2 + (3)^2$$

$$= 1296 + 9 = 1305$$

$$\therefore AD = \sqrt{1305} = 36.12$$

$$\therefore AD = BC = 36.12 \text{ padas.}$$



SAUTRĀMANIVEDI

वेदितृतीये यजेतेति सौत्रामणिकी वेदिमश्युपदिशन्ति॥ (BSS 1.85)

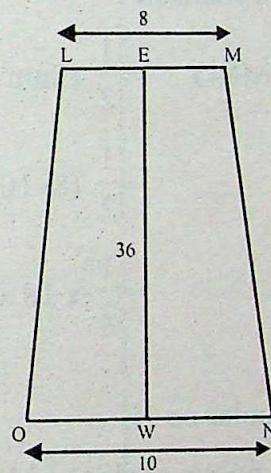
"He is to sacrifice on the third part of the vedi" thus it explains the formation of vedi of the Sautramani sacrifice.

महावेदेस्तृतीयेन समचतुरस्कृताया अष्टादशपदा पाश्वर्वमानी भवति॥ (BSS 1.86)

The side of that area which is formed into a square with the third part of the Mahavedi is eighteen padas long.

तस्यै दीर्घकरण्यामेकतोडणिमत्करण्यां च याथाकामीति ॥ (BSS 1.87)

The meaning of the sutra is : instead of giving to the Sautrāmanivedi the form of a square we may give to it the form of an "oblong which is shorter on one side" and the area of which is equal to the third part of the Mahavedi [17]..



Description :

The Sautrāmanivedi is constructed exactly in the same manner as the Mahavedi. Measures prescribed for the Mahavedi are 36 padas for prāci (east-west line), 30 padas for western side, 24 padas for eastern side all these lengths were devided by three then we obtain 12 padas for prāci, 10 padas for western side and 8 padas for estern side. The area of figure formed is one ninth of the area of Mahavedi which is 108 units of area; Multiplying it by 3, we get 324 units of area which is equal to the area of Sautramani vedi and dimensions of Sautramani vedi are as follows 8,10 and 36 padas [4].

Measurment of area of Sautramani vedi by Sulba Sutra method:

$$AB = 12 \text{ Padas}$$

$$AE = EG = GD = 9 \text{ padas}$$

$$\therefore AF = FG = HD = 108 \text{ sq. padas}$$

$$\therefore AC = 324 \text{ sq. padas.}$$

$$\text{Area of rectangle ABFE} = 12 \times 9 = 108 \text{ sq. padas}$$

$$\therefore \text{Area of rectangle ABCD} = 12 \times 27 = 324 \text{ sq. padas}$$

$$\text{Area of Sautramanivedi LMNO} = \text{Area of rectangle ABCD} = 324 \text{ sq. padas [4].}$$

Varification of area by modern method :

$$\begin{aligned} \text{Area of Sautramanivedi LMNO} &= \frac{1}{2} (\text{Sum of parallel sides}) \times \text{height} \\ &= \frac{1}{2} (8+10) \times 36 \\ &= \frac{1}{2} \times 18 \times 36 = 9 \times 36 = 324 \text{ sq. padas} \end{aligned}$$

Length of non-parallel sides:

From fig., MP is perpendicular on N.

Since E and W are middle points of sides LM and ON

$$\therefore LE = EM = 4 \text{ and } OW = WN = 5$$

$$\therefore PN = WN - WP$$

$$= 5 - 4 = 1 \quad (\therefore WP = EM)$$

Since EW = 36

$$\therefore MP = 36$$

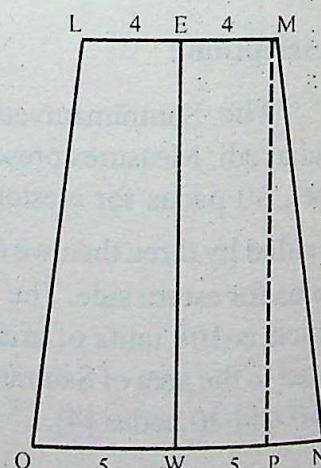
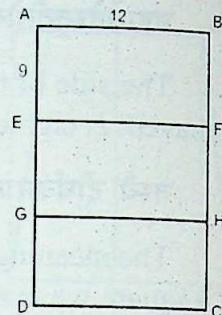
In triangle MPN, using Sulbakra's theorem

$$MN^2 = MP^2 - PN^2$$

$$= (36)^2 + (1)^2$$

$$\therefore MN = \sqrt{1297} = 36.01 \text{ padas}$$

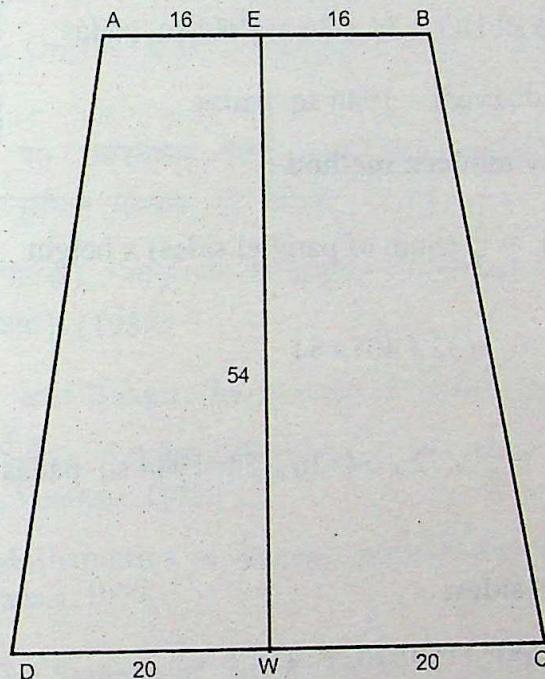
$$\therefore MN = LO = 36.01 \text{ padas}$$



ASVAMEDHA VEDI

द्विस्तावा वेदिर्भवतीत्यश्वमेधे विज्ञायते॥ (ASS 5.36)

Area of the altar for Aśvamedha is double the area of Saumikivedi (Mahavedi) i.e. 1944 sp. pada [4].



Description :

Because area of Mahavedi is 972 sq. padas therefore area of Aśvamedha vedi will be 1944 sq. padas. Shape of this vedi is the same as that of Mahavedi. Accordingly prāci of Aśvamedha vedi is 54, western side is 40 and eastern side is 32 padas [4].

Now in figure ABCD is Aśvamedhavedi.

Measurement of area of Asvamedhavedi by Sulba Sutra Method :

Draw two perpendicular one from B on DC and second from D on BA. Let perpendiculars meet on WC at F and on BA at G.

Area of rectangle BFDG = Area of Asvamedhavedi ABCD

Since E and W are middle points of the sides AB and DC.

$$\therefore AE = EB = 16 \text{ and } DW = WC = 20$$

$$\text{also } GD = EW = BF = 54$$

$$\therefore DF = DW + WF = 20 + 16 = 36$$

$$\therefore \text{Area of rectangle BFDG} = 54 \times 36 = 1944 \text{ sq. padas}$$

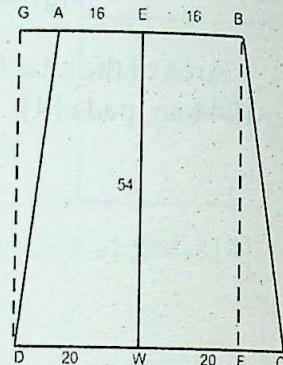
Hence area of Asvamedhavedi = 1944 sq. padas.

Varification of area by modern method :

$$\text{Area of Asvamedhavedi} = \frac{1}{2} (\text{Sum of parallel sides}) \times \text{height}$$

$$= \frac{1}{2} (32 + 40) \times 54$$

$$= \frac{1}{2} \times 72 \times 54 = 36 \times 54 = 1944 \text{ sq. padas}$$



Length of non-parallel sides:

$$\text{from fig., } FC = WC - WF = 20 - 16 = 4$$

$$\text{and } BF = 54$$

In triangle BFC, using Sulbakra's theorem

$$BC^2 = NF^2 + FC^2$$

$$BC^2 = (54)^2 + (4)^2$$

$$= 2916 + 16 = 2932$$

$$\therefore BC = \sqrt{2932} = 54.15$$

$$\therefore BC = AD = 54.15 \text{ padas.}$$

In this paper the dimensions of Mahavedi, Sautrāmanivedi and Aśvamedhavedi and their shapes are described. Their areas have been calculated by Śulba Sūtra method and modern method. In each case area comes out to be equal, showing the exactness of the geometrical methods described in Śulba Sūtra.

REFERENCES

1. Archaeological Survey of India, Annual Report, (ASI, AR), 1920-21, P.17.
2. Afzal Ahmad : On the π of Aryabhata I, Ganita Bharati (GB), 3 (1981) 83-85.
3. सत्यप्रकाश एवं प० रामस्वरूप शर्मा (सम्पादक), आपस्तम्ब शुल्बसूत्रम्, प्राचीन वैज्ञानिकाध्ययन अनुसंधान संस्थान, नई दिल्ली, (1938)
4. दामोदर ज्ञा (संपादक), आपस्तम्ब-शुल्बसूत्रम्, विश्वेश्वरानन्द वैदिक शोध संस्थान, होशियारपुरम् (पंजाब) (1988).
5. Bodhāyanācār and Bhagiratha Prasada Tripathi 'Vagisasatri' (Ch. Ed.): Baudhāyana Śulba Sutrām, Research Institute, Sampurnanad Sanskrit Visvavidyalaya, Varanasi, 1979.
6. A.K. Bag, : Mathematics in Ancient and Medieval India, Chaukhamba Orientailia, Varanasi, 1979.
7. B. Datta, : The Science of Śulba, A study in Early Hindu Geometry; Calcutta University, 1932.
8. B. Datta and A.N. Singh : History of Hindu Mathematics (Part I and II), Asia Publishing House, Bombay, 1962.
9. R.C. Gupta, : Vedic Mathematics from the Śulba Sūtras; J. of Mathematics Education 9 (2) (1989) 1-10.
10. Nidhi Handa, Kātyāyana Śulba Sūtras and Modern Mathematical Interpretations of its Sutras, Ph.D. Thesis, Gurukul Kangri Uni, Haridwar.
11. Pt. Lashan Lal Jha. (Ed.) : Lilāvati of Bhāskarācārya II, The Chowkhamba Vidya Bhawan, Varanasi Chowk 1961.
12. S.D. Khadikar (Ed.): Kātyāyana Śulba Sūtras, Vaidika Samsodhana Mandoler, Poona, 1974.

13. Vinod Mishra: A Study of Vedic Geometry and its Relevance to Science and Technology, Ph.D. Thesis, Gurukul Kangri Uni, Haridwar (1977).
14. S.N.Sen: A Bibliography of Sanskrit works on Astronomy and Mathematics part I, National Institute of Sciences of India, New Delhi, 1966.
15. V.K. Sharma and Yogita Bana : Some Geometrical Construction form Baudhayana Sulba Sutra, J. Natural & Phys. Sci. 18 (1) (2004) 57-67.
16. Sharma, V.K. and Yogita Bana : Karani and related Construction from Baudha \bar{y} ana Šulba Sūtras, ACCST Research Journal, 11 (2004) 4-10
17. D.N.Yajvan and G. Thibout : Baudhayana Sulba Sutram, The Research Institute Ancient Scientific Studies, New Dehli, 1968.

KINETICS AND MECHANISM OF MN (II) CATALYSED PERIODATE OXIDATION OF O-CHLORO ANILINE IN AETONE-WATER MEDIUM

R.D. KAUSHIK*, YOGESH C. NAINWAL, VIRENDRA SINGH AND A.K. CHAUBEY

(Received 10.08.2004)

ABSTRACT

Kinetic-mechanistic studies made for periodate oxidation of o-chloroaniline (OCA) in acetone-water medium, have been used for derivation of rate law and proposing mechanism which satisfy various observations like first order in each reactant and catalyst, stoichiometry (1 mol OCA : 2 mol periodate), kinetic parameters ($E_a = 6.5$ K cals/mol; $A = 1.45 \times 10^5$ lit/mol/sec; $\Delta S^\ddagger = -38.0$ EU; $\Delta F^\ddagger = 17.72$ K Cals/mol; $\Delta H^\ddagger = 5.9$ K Cals/mol), main product identified (2-chloro-1,4 benzoquinone), and the effect of pH, dielectric constant, and ionic strength on reaction rate.

Key words : Kinetics and mechanism, Mn (II) catalyst, Periodate oxidation, o-chloroaniline, 2-chloro-1,4-benzoquinone.

INTRODUCTION

Reports on kinetic-mechanistic studies on the Mn (II) catalysed periodate oxidation of aromatic amines are very few[1-4]. In continuation to our earlier communications on uncatalysed oxidations of such type [5-7], present paper deals with the kinetic studies made on Mn (II) catalysed periodate oxidation of OCA in acetone-water medium.

MATERIALS AND METHODS

Chemicals of E.Merck/CDH A.R. grade were used after distillation/recrystallization. Triply distilled water was used for preparation of the solutions. The progress of the reaction was followed spectrophotometrically [5] by recording the absorbance at 475 nm, i.e. the λ_{max} of reaction mixture in the duration in which the λ_{max} did not change. The pH was maintained at 5.0 by using Thiel, Schultz and Koch buffer [6,7]

in all kinetic runs except when the effect of pH was studied, Plane mirror method and Guggenheim's method were used for evaluation of initial rates $[(dA/dt)_i]$ and pseudo first order rate constant k_1 or second order rate constant k_2 respectively. NaCl solutions were used for maintaining the ionic strength (μ) in the kinetic runs.

RESULTS AND DISCUSSION

1 mol OCA consumed 2 moles of periodate as determined iodometrically. The data (table-1) indicated 2nd order for the reaction, being first order in each reactant. Linear relation between concentration of the reactants and rate supported the 2nd order kinetics. In pseudo first order conditions (table-2), the $[(dA/dt)]^{-1}$ or k_1^{-1} vs $[S]^{-1}$ Plots were linear with almost negligible intercept, suggesting the Michaelis-Menten type kinetics being followed with respect to both reactants with the possibility of formation of a fast decaying intermediate complex between reactants [10,11]. Data in table-3 established the 1st order in catalyst. Rate-pH profile showed a maxima at pH=5.0 (table-4). A linear relation between $\log (dA/dt)_i$ or $\log k_2$ and $1/D$ with negative slope (where D is the dielectric constant of the medium) and a primary linear type plot between $\log (dA/dt)_i$ or $\log k_2$ vs ionic strength (μ) that were obtained from the data in table-5, indicated an ion-dipole interaction in this reaction. Arrhenius plot was made between $30 \pm 0.1^\circ\text{C}$ to $45 \pm 0.1^\circ\text{C}$ and the values of different thermodynamic parameters evaluated taking $[\text{OCA}] = 1.0 \times 10^{-3}\text{M}$, $[\text{NaIO}_4] = 1.0 \times 10^{-4}\text{M}$, $[\text{Mn}^{++}] = 8.0 \times 10^{-6}\text{M}$ and acetone = 10.0% (v/v), are $E_a = 6.5\text{ K cals/mol}$; $A = 1.45 \times 10^5\text{ lit/mol/sec}$; $\Delta S^* = -38.0\text{ EU}$; $\Delta F^* = 17.72\text{ K cals/mol}$; $\Delta H^* = 5.9\text{ K cals/mol}$. Low value of energy of activation and high frequency factor are characteristic of a bimolecular reaction in the solution in the which the reacting species are larger in size. A large negative value of ΔS^* suggests the formation of strongly solvated, charged and rigid transition state.

Table-1. Determination of order w.r.t. reactants.

$\lambda_{\max} = 475 \text{ nm}$; pH = 5.0; Acetone = 10.0% (v/v); Temp = $30 \pm 0.1^\circ\text{C}$;											
$[\text{Mn}^{++}] = 8.0 \times 10^{-6} \text{ M}$ ($6 \times 10^{-6} \text{ M}$ for variation of $[\text{NaIO}_4]$)											
$[\text{OCA}] \times 10^4 \text{ M}$	2.0	4.0	6.0	8.0	10.0	10.0	10.0	10.0	10.0	10.0	10.0
$[\text{NaIO}_4] \times 10^3 \text{ M}$	1.0	1.0	1.0	1.0	1.0	0.1	0.2	0.3	0.4	0.8	1.0
$(dA/dt) \times 10^2 (\text{min}^{-1})$	3.0	5.6	8.0	10.3	12.4	7.0	14.0	27.0	40.0	54.0	64.0

Table-2. $\lambda_{\max} = 475 \text{ nm}$; pH = 5.0; Acetone = 10.0% (v/v); Temp = $30.0 \pm 0.1^\circ\text{C}$;
 $[\text{Mn}^{++}] = 8.0 \times 10^{-6} \text{ M}$ ($6 \times 10^{-6} \text{ M}$ for variation of $[\text{NaIO}_4]$).

$[\text{OCA}] \times 10^3 \text{ M}$	1.0	1.2	1.4	1.6	1.8	0.1	0.1	0.1	0.1	0.1	0.1
$[\text{NaIO}_4] \times 10^3 \text{ M}$	0.10	0.10	0.10	0.10	0.10	1.0	2.0	3.0	4.0	5.0	6.0
$(dA/dt) \times 10^3 (\text{min}^{-1})$	6.5	14.0	21.0	27.5	32.5	7.0	15.0	22.0	26.0	34.0	40.0
$k \times 10^3 (\text{sec}^{-1})$	1.73	2.01	2.30	2.50	2.78	1.54	2.30	3.07	3.45	3.84	4.22

Table-3. Determination of order w.r.t. Mn^{++}

$[\text{NaIO}_4] = 1.0 \times 10^{-4} \text{ M}$; $[\text{OCA}] = 1.0 \times 10^{-3} \text{ M}$; Temp. = $30.0 \pm 0.1^\circ\text{C}$;											
Acetone = 5.0% (v/v); $\lambda_{\max} = 475 \text{ nm}$											
$[\text{Mn}^{++}] \times 10^{-6} \text{ M}$	4.0	6.0	8.0	10.0	12.0	14.0					
$(dA/dt) \times 10^3 (\text{min}^{-1})$	6.5	10.0	13.0	16.0	19.0	22.0					

Table-4. Rate-pH profile.

$[\text{OCA}] = 1.0 \times 10^{-3} \text{ M}$; $[\text{NaIO}_4] = 1.0 \times 10^{-4} \text{ M}$; $\lambda_{\max} = 475 \text{ nm}$; Acetone = 5.0% (v/v);											
Temp. = $30.0 \pm 0.1^\circ\text{C}$; $[\text{Mn}^{++}] = 8.0 \times 10^{-6} \text{ M}$											
pH	3.5	4.0	4.5	5.0	5.5	6.0	6.5	7.0	7.5		
$(dA/dt) \times 10^3 (\text{min}^{-1})$	5.0	5.25	6.5	8.0	5.5	5.0	4.5	4.0	3.0		

Table-5 Effect of D and μ on reaction rate.

[OCA] = 1.0×10^{-3} M; $[\text{NaIO}_4] = 1.0 \times 10^{-4}$ M; $\lambda_{\text{max}} = 475$ nm; * Acetone = 5.0% (v/v);

Temp. = $30.0 \pm 0.1^\circ\text{C}$; $[\text{Mn}^{++}] = 8.0 \times 10^{-6}$ M.

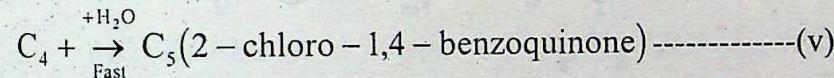
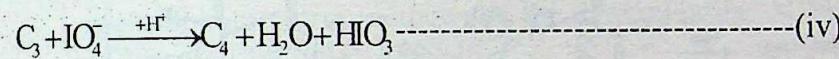
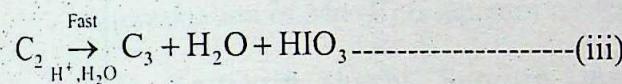
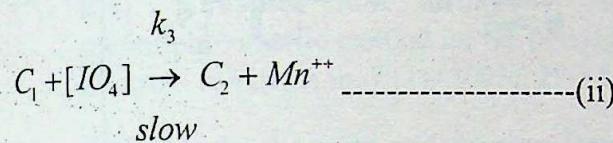
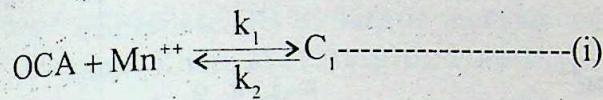
D	72.4	71.0	69.7	68.4	--	--	--	--
* $\mu \times 10^2$	--	--	--	--	10.0	12.3	14.1	17.3
$(dA/dt)_i \times 10^3$ (min ⁻¹)	9.0	8.0	7.0	6.0	6.5	7.0	7.5	8.25
$k_1 \times 10^3$ (Sec ⁻¹)	3.65	2.40	1.73	1.34	1.54	1.60	1.63	1.73
$k_2 (1. \text{ mol}^{-1} \cdot \text{Sec}^{-1})$	3.65	2.40	1.73	1.34	1.54	1.60	1.63	1.73

The increase in the rate from pH 3.5 to 5.0 may be due to the decrease in the protonation of OCA from pH 3.5 to 5.0 (table-4) which makes greater concentration of OCA available for the reaction. This leads to the assumption that unprotonated OCA is the reactive species in present case. Second part of this profile suggests that the periodate monoanion $[\text{IO}_4^-]$ is the reactive species of periodate [9] whose concentration goes on decreasing with increase in pH beyond the value 5.0 [9] decreasing thereby the rate of reaction beyond pH 5.0.

Based on these results, the mechanism may be proposed (Chart) involving the lone pair of electrons on nitrogen atom of OCA for the co-ordinate bond formation between OCA and Mn^{++} species in a reversible step to from complex C_1 [step (i)] which, in turn, interacts with IO_4^- in slow and rate determining step (ii) to give C_2 . C_2 Changes by fast hydrolysis in to C_3 . The formation of a charged intermediate complex C_2 by the attack of IO_4^- on the nitrogen of anilino group and stabilization of positive charge on nitrogen of this group, have also been observed and supported by LFER studies for the uncatalysed periodate oxidation of few aromatic amines [8]. In addition, a high negative value of entropy of activation and the effect of dielectric constant on the reaction rate support the involvement of solvation effects in this reaction.

It should also be noted that the initial part of the reaction is significant in the present case and the second molecule of IO_4^- reacting later to give C_4 is not significant. C_4 changes by fast hydrolysis to give C_5 i.e. the main product of reaction that has been isolated and separated by employing the method reported earlier in case of other aromatic amines [6,7]. It was characterized as 2-chloro-1,4-benzoquinone on the basis of test for quinone [4], M.P. [12], UV-VIS, IR and H-nmr spectra (showing

characteristic bands for C-C1 [13] and benzoquinones [14]). The overall process may be represented as follows:



On applying steady state treatment to C_1 , the rate law in terms of rate of loss of $[\text{IO}_4^-]$ may be derived as follows:

$$\text{Rate of loss of } [\text{IO}_4^-] \text{ or } -d[\text{IO}_4^-]/dt = k_3 [\text{C}_1] [\text{IO}_4^-]$$

$$= \text{Rate of loss of } \text{C}_1 \text{ or } -d[\text{C}_1]/dt \quad \text{(i)}$$

$$\text{Rate of formation of } \text{C}_1 = +d[\text{C}_1]/dt = k_1 [\text{OCA}] [\text{Mn}^{++}] - k_2 [\text{C}_1]$$

$$\text{At steady state, } -d[\text{C}_1]/dt = +d[\text{C}_1]/dt$$

$$\text{Therefore, } k_3 [\text{C}_1] [\text{IO}_4^-] = k_1 [\text{OCA}] [\text{Mn}^{++}] = k_2 [\text{C}_1]$$

$$\text{or } [\text{C}_1] \{k_3 [\text{IO}_4^-] + k_2\} = k_1 [\text{OCA}] [\text{Mn}^{++}]$$

$$\text{or } [\text{C}_1] = \frac{k_1 [\text{OCA}] [\text{Mn}^{++}]}{k_3 [\text{IO}_4^-] + k_2} \quad \text{(2)}$$

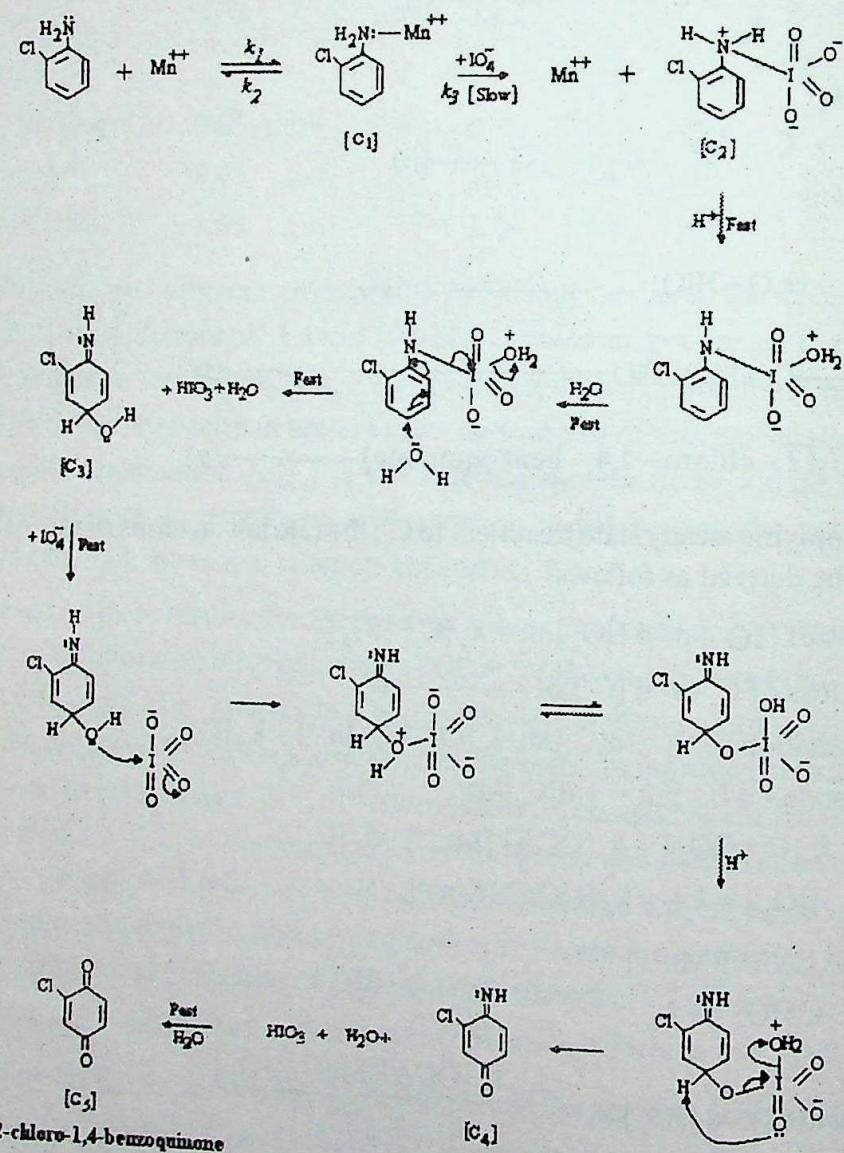
$$\text{From (1) and (2), } -d[\text{IO}_4^-]/dt = \frac{k_3 \cdot k_1 [\text{OCA}] [\text{Mn}^{++}] [\text{IO}_4^-]}{k_2 + k_3 [\text{IO}_4^-]} \quad \text{(3)}$$

Since step (ii) is slow and rate determining step, hence $k_3 [\text{IO}_4^-] \ll k_2$ may be assumed. Therefore, the rate law (3), may be written as rate law (4) which explains all the kinetic results.

$$-d[\text{IO}_4^-]/dt = k_{\text{obs}} [\text{OCA}] [\text{Mn}^{++}] [\text{IO}_4^-] \quad \text{(4)}$$

$$\text{where, } k_{\text{obs}} = k_3 k_1 / k_2$$

CHART



REFERENCES

1. I.F. Dolmanova, V.P. Poddubienko, and V.M. Peshkova: Determination of manganese (II) by kinetic methods using the oxidation of p-phenetidine, *Zh. Anal. Khim.* 25 (11) (1970), 2146-2150.
2. R.D. Kaushik, A.K. Chaubey and R.P. Singh: A new kinetic-spectrophotometric method for the determination of Mn (II) in water, *Indian J. Environ. & Ecoplann.* 7 (1) (2003), 29-34.
3. R.D. Kaushik, Amrita and Sumitra Devi : An improved method for nanogram determination of Mn (II) in aqueous medium, *J. Curr. Sci.* 3 (1) (2003), 197-202.
4. R.D. Kaushik, Shashi, Sumitra Devi and R.P. Singh: Simple kinetic-spectrophotometric method for the determination of Mn (II) in water, *Asian J.Chem* 16 (2) (2004), 837-840.
5. R.D. Kaushik et al.: *Asian J.Chem.*, 11 (2) 633 (1999); 12 (4), 1123, 1129, 1229 (2000).
6. R.D. Kaushik, R.P. Singh and Shashi: Kinetic-mechanistic study of periodate oxidation of p-chloroaniline, *Asian J. Chem* 15 (3 & 4) (2003), 1655-1658.
7. R.D. Kaushik, A.K. Chaubey and P.K. Garg: Kinetics and mechanism of periodate oxidation of p-phenetidine, *Asian J.Chem.* 15 (3 & 4) (2003), 1485-1490.
8. S.P. Srivastava, V.K. Gupta M.C. Jain, M.N. Ansari and R.D. Kaushik: Thermodynamic and LFER studies for the oxidation of anilines by periodate ion, *Thermo Chimica Acta*, 68 (1983), 27-33.
9. V.K. Pavolva, Ya. S. Sevchenko and K.B. Yatsmiriskii: Kinetics and mechanism of oxidation of diethylaniline by periodate, *Zh.Fiz. Khim.*, 44 (3) (1970), 658-63.
10. R.D. Kaushik, Shashi, Amrita and Sumitra Devi: Kinetics and mechanism of Mn (II) catalysed periodate oxidation of 4-chloro-2-methylaniline, *Asian J. Chem.* 16 (2) (2004), 818-822.
11. N. Nalwaya, A. Jain and B.L. Hiran : Kinetics of oxidation of glycine by pyridinium bromo chromate in acetic acid medium, *J. Indian Chem. Soc.* 79 (7) (2002), 587-589.

12. J. Buckingham (Ed.): Dictionary of Organic compounds, 5th edn., Vol. I, Chapman & Hall, New York, 1982.
13. J.R. Dyer : Application of absorption spectroscopy to Organic compounds, Prentice-Hall of India, New Delhi, 1984 p 38.
14. R.M. Silverstein : Spectrometric identification of organic compounds, 5th ed., John Wiley and sons. Inc. New York 1991.

STRONG SUMMABILITY OF FUNCTIONS BASED ON (D, α, β) (C, ℓ, m) SUMMABILITY METHODS

B.P. MISHRA* & S.K. SINGH*

(Received 10.08.2004)

ABSTRACT

Singh [7] defined strong summability methods. Here, we have defined summability of function based on (D, α, β) , (C, ℓ, m) summability methods and investigated some of their properties.

Key words and Phrases: Strong summability, function, (D, α, β) , (C, ℓ, m) summability methods,

Lebesgue integral, Holder's inequality and Minkowski's inequality.

Mathematics subject classifications (2000): 40F05.

SOME NOTATIONS AND DEFINITIONS

Let $s(x)$ be any function which is Lebesgue integrable in $(0, x)$ for all finite $x > 0$ and that $s(x)$

is bounded in some right hand neighbourhood of the origin. Integrals of the form \int_0^∞ are

throughout to be taken as $\lim_{n \rightarrow \infty} \int_0^x$, \int_0^x being taken as a Lebesgue integral. We write

$$\partial_{\ell, m}(x) = \begin{cases} S(x) & l = 0 \\ \left\{ \frac{\Gamma(\ell + m + 1)}{\Gamma(\ell)\Gamma(m + 1)} \right\} \frac{1}{X^{\ell+m}} \int_0^x (x - y)^{\ell-1} y^m s(y) dy, & \ell > 0, m > -1 \end{cases} \quad (1.1)$$

We also write

$$g^{\alpha, \beta}(y) = \left\{ \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} \right\} y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1}}{(x + y)^{\alpha+\beta+1}} s(x) dx, \quad (1.2)$$

if this exists. We also write

$$U_{\alpha, \beta, \ell, m}(y) = \left\{ \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} \right\} y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1}}{(x + y)^{\alpha+\beta+1}} \partial_{\ell, m}(x) dx, \quad (1.3)$$

if this exists, where $\ell > 0, m > -1, y > 0, \beta > -1, \alpha > 0$.

With the usual terminology, we say that $s(x)$ is summable

(i) (C, ℓ, m) to the sum s , if $\partial_{\ell, m(x)}(x)$ exists, and tends to a limit s as $x \rightarrow \infty$,

(ii) $[C, \ell, m]_p$ or strongly summable (C, ℓ, m) with index p to the sum s , where

$\ell \geq 1, m \geq -1, p \geq 1$, if

$$\int_0^x |\partial_{\ell-1, m}(y) - s|^p dy = o(x) \text{ as } x \rightarrow \infty \quad (1.4)$$

(iii) (D, α, β) to the sum s , if $g^{\alpha, \beta}(y)$ tends to s as $y \rightarrow \infty$,

(iv) $(D, \alpha, \beta) (C, \ell, m)$ to the sum s , if $U_{\alpha, \beta, \ell, m}(y)$ tends to s as $y \rightarrow \infty$.

Definition (ii) & (iii) are due respectively to Mishra and Srivastava [5] and Kwee [3], while (iv) is due to Mishra and Singh [6].

Let

$$s(x) = \int_0^x a(y) dy, \quad (1.5)$$

where $a(y)$ is Lebesgue integrable in $(0, x)$ for any finite $x > 0$. Since $U_{\alpha, \beta, \ell-1, m}(y)$ converges for all $y > 0$, then it is easy to justify the differentiation under integral sign (\int) and that

$$\begin{aligned} y \frac{d}{dy} U_{\alpha, \beta, \ell, m}(y) &= -(\ell + m)[U_{\alpha, \beta, \ell, m}(y) - U_{\alpha, \beta, \ell-1, m}(y)] \\ &= B_{\alpha, \beta, \ell, m}(y) \quad (\ell \geq 1), \end{aligned} \quad (1.6)$$

where

$$B_{\alpha, \beta, \ell, m}(y) = \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1}}{(x+y)^{\alpha+\beta+1}} \tau_{\ell, m}(y) dy.$$

Also we know that

$$y \frac{d}{dy} \partial_{\ell,m}(y) = (\ell + m)[\partial_{\ell-1,m}(y) - \partial_{\ell,m}(y)] = \tau_{\ell,m}(y) \quad (1.7)$$

$$= \frac{\Gamma(\ell + m + 1)}{\Gamma(\ell)\Gamma(m + 1)} \frac{1}{y^{\ell+m}} \int_0^y (y - v)^{\alpha-1} v^m v a(v) dv.$$

Before giving the justification of (1.6) we first make some remarks on (1.7). If we do not assume (1.5), then $\tau_{\ell,m}(y)$ has not been defined, so that the second equation (1.7) is meaningless. However the first equality holds every where if $\ell \geq 2$ and almost every where if $2 > \ell \geq 1$. if we assume (1.5), then both the equations in (1.7) hold every where for $\ell \geq 1$, $y \geq 1$, the middle term in (1.7) is not defined. However, the equality

$$y \frac{d}{dy} \partial_{\ell,m}(y) = \tau_{\ell,m}(y)$$

continues to hold almost every where for $1 > \ell > 0$.

In order to get a valid proof of (1.6), under the assumption that the integral defining $U_{\alpha,\beta,\ell-1,m}(y)$ converges for all $y > 0$, that it is necessary and sufficient for the convergence of (1.3), that

$$\int_1^\infty \frac{\partial_{\ell-1,m}(x)}{x^{\beta+2}} dx \text{ should converge} \quad (1.8)$$

We write

$$R_{\ell-1,m}(y) = \int_y^\infty \frac{\partial_{\ell-1,m}(x)}{x^{\beta+1}} dx$$

which is bounded for $y \geq 0$ and tends to zero as $y \rightarrow \infty$.

Now

$$\partial_{\ell,m}(u) = \frac{\ell + m}{u^{\ell+m}} \int_0^u v^{\ell+m-1} \partial_{\ell-1,m}(v) dv$$

$$\begin{aligned}
 &= -\frac{(\ell+m)}{u^{\ell+m}} \int_0^u v^{\ell+m-1} v^{\beta+1} dR_{\ell-1,m}(v) \\
 &= -\frac{(\ell+m)}{u^{\ell+m}} \int_0^u v^{\ell+m+\beta+1} dR_{\ell-1,m}(v) \tag{1.9}
 \end{aligned}$$

Since $R_{\ell-1,m}(v) \rightarrow \infty$, therefore given $\epsilon > 0$, there is number u_0 such that $|R_{\ell-1,m}(v)| < \epsilon$ for $v > u_0$, we also note that $|R_{\ell-1,m}(v)| = O(1)$ for $v < u_0$. Hence we see from (1.9), on integration by parts, that

$$\begin{aligned}
 \partial_{\ell,m}(u) &= -\frac{\ell+m}{u^{\ell+m}} \{v^{\ell+m+\beta+1} R_{\ell-1,m}(v)\} + \frac{\ell+m}{u^{\ell+m}} \int_0^u \frac{d}{dv} (v^{\ell+m+\beta+1}) R_{\ell-1,m}(v) dv \\
 &= -\frac{\ell+m}{u^{\ell+m}} \int_0^u \frac{d}{dv} (v^{\ell+m+\beta+1}) R_{\ell-1,m}(v) dv + \frac{\ell+m}{u^{\ell+m}} \int_{u'}^u \frac{d}{dv} (v^{\ell+m+\beta+1}) R_{\ell-1,m}(v) dv \\
 &= O(u^{\beta+1}) + o(u^{\beta+1}) \\
 &= o(u^{\beta+1}), \text{ if } u \rightarrow \infty
 \end{aligned}$$

Hence $\partial_{\ell,m}(u) = O(u^{\beta+1})$, as $u \rightarrow \infty$ (1.10)

$$\text{Let } H_{\alpha,\beta,\ell,m}(y) = \int_0^\infty \frac{x^{\alpha-1}}{(x+y)^{\alpha+\beta+1}} \partial_{\ell,m}(x) dx$$

$$H_{\alpha,\beta,\ell,m}(\eta+y) = \int_0^\infty \frac{x^{\alpha-1}}{(x+\eta+y)^{\alpha+\beta+1}} \partial_{\ell,m}(x) dx$$

Let $y > y_0$ to be fixed. By Lagrange's formula, for the remainder in Taylor's series, we have, for $\eta > -y$, and for all $x \geq 0$

$$\frac{1}{\eta} \{H_{\alpha,\beta,\ell,m}(\eta+y) - H_{\alpha,\beta,\ell,m}(y)\} = \frac{1}{\eta} \int_0^\infty x^{\alpha-1} \partial_{\ell,m}(x) \left\{ \frac{1}{(x+n+y)^{\alpha+\beta+1}} - \frac{1}{(x+y)^{\alpha+\beta+1}} \right\} dx$$

By Lagrange's formula, we have

$$\frac{1}{\eta} \left\{ \frac{1}{(x + \eta + y)^{\alpha + \beta + 1}} - \frac{1}{(x + y)^{\alpha + \beta + 1}} \right\} = - \frac{(\alpha + \beta + 1)}{(x + y)^{\alpha + \beta + 2}} + \frac{\eta}{2} \frac{(\alpha + \beta + 1)(\alpha + \beta + 2)}{(\xi + y)^{\alpha + \beta + 3}}$$

for some ξ between $(x+y)$ and $(x+\eta+y)$ (hence ξ depends on x, y, η). Now choose a fixed y' with $0 < y' < y$. Then, for $\eta > -(y - y')$ (and thus, in particular, for all sufficiently small η), we have $\xi \geq y$,

$$\begin{aligned} & \left| \frac{1}{\eta} \{H_{\alpha, \beta, \ell, m}(\eta + y) - H_{\alpha, \beta, \ell, m}(y)\} + \int_0^\infty \frac{(\alpha + \beta + 1)x^{\alpha-1}}{(x + y)^{\alpha + \beta + 2}} \partial_{\ell, m}(x) dx \right| \\ & \leq \frac{|\eta|}{2} \left| \int_0^\infty x^{\alpha-1} \partial_{\ell, m}(x) \frac{(\alpha + \beta + 1)(\alpha + \beta + 2)}{(x + y)^{\alpha + \beta + 3}} dx \right| \\ & \leq o(|\eta|). \end{aligned}$$

Since the integral on the right hand side converges. Hence $H_{\alpha, \beta, \ell, m}(y)$ is differentiable, and

$$\begin{aligned} \frac{d}{dy} H_{\alpha, \beta, \ell, m}(y) &= \lim_{\eta \rightarrow 0} \frac{1}{\eta} [H_{\alpha, \beta, \ell, m}(\eta + y) - H_{\alpha, \beta, \ell, m}(y)] \\ &= -(\alpha + \beta + 1) \int_0^\infty \frac{x^{\alpha-1}}{(x + y)^{\alpha + \beta + 2}} \partial_{\ell, m}(x) dx. \end{aligned} \quad (1.12)$$

We know that

$$U_{\alpha, \beta, \ell, m}(y) = \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1}}{(x + y)^{\alpha + \beta + 1}} \partial_{\ell, m}(x) dx$$

Differentiating this equation, we have

$$\begin{aligned} \frac{d}{dy} U_{\alpha, \beta, \ell, m}(y) &= \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} [(\beta + 1)y^\beta \int_0^\infty \frac{x^{\alpha-1}}{(x + y)^{\alpha + \beta + 1}} \partial_{\ell, m}(x) dx \\ &\quad - y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1}(\alpha + \beta + 1)}{(x + y)^{\alpha + \beta + 2}} \partial_{\ell, m}(x) dx] \end{aligned}$$

$$\begin{aligned}
 &= \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} \left[y^\beta \int_0^\infty \frac{x^{\alpha-1}}{(x+y)^{\alpha+\beta+1}} \left\{ (\beta+1) - \frac{y(\alpha+\beta+1)}{(x+y)} \right\} \partial_{\ell,m}(x) dx \right] \\
 &= \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} \left[y^\beta \int_0^\infty \frac{x^{\alpha-1}}{(x+y)^{\alpha+\beta+1}} \left\{ (\beta+1) - \frac{y(\alpha+\beta+1)}{(x+y)} \right\} \partial_{\ell,m}(x) dx \right] \\
 &= \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^\beta \int_0^\infty \frac{x^{\alpha-1}}{(x+y)^{\alpha+\beta+2}} \left\{ x(\beta+1) - y\alpha \right\} \partial_{\ell,m}(x) dx.
 \end{aligned}$$

Hence

$$y \frac{d}{dy} U_{\alpha, \beta, \ell, m}(y) = \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1}}{(x+y)^{\alpha+\beta+2}} \left\{ x(\beta+1) - y\alpha \right\} \partial_{\ell,m}(x) dx. \quad (1.13)$$

Now, from the middle term of (1.6), we have

$$\begin{aligned}
 -(\ell+m)[U_{\alpha, \beta, \ell, m}(y) - U_{\alpha, \beta, \ell-1, m}(y)] &= -(\ell+m) \left[\frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1}}{(x+y)^{\alpha+\beta+1}} \partial_{\ell,m}(x) dx \right. \\
 &\quad \left. - \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1}}{(x+y)^{\alpha+\beta+1}} \left\{ \partial_{\ell-1,m}(x) dx \right\} \right] \quad (1.14) \\
 &= -(\ell+m) \left[\frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1}}{(x+y)^{\alpha+\beta+1}} \left\{ \partial_{\ell,m}(x) - \partial_{\ell-1,m}(x) \right\} dx \right] \\
 &= -(\ell+m) \left[\frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1}}{(x+y)^{\alpha+\beta+1}} \left\{ -\frac{x}{\ell+m} \frac{d}{dx} \partial_{\ell,m}(x) \right\} dx \right] \\
 &= \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^{\beta+1} \int_0^\infty \frac{x^\alpha}{(x+y)^{\alpha+\beta+1}} d\partial_{\ell,m}(x) dx
 \end{aligned}$$

Integrating by parts, we see that the above expression becomes

$$\begin{aligned}
 &= \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^{\beta+1} \left[\left\{ \frac{x^\alpha}{(x+y)^{\alpha+\beta+1}} \partial_{\ell,m}(x) \right\}_0^\infty - \right. \\
 &\quad \left. \int_0^\infty \frac{(x+y)^{\alpha+\beta+1} \alpha x^{\alpha-1} - x^\alpha (\alpha + \beta + 1)(x+y)^{\alpha+\beta}}{(x+y)^{2(\alpha+\beta+1)}} \partial_{\ell,m}(x) dx \right] \\
 &= - \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1} \{\alpha(x+y) - x(\alpha + \beta + 1)\}}{(x+y)^{\alpha+\beta+2}} \partial_{\ell,m}(x) dx \\
 &= - \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1}}{(x+y)^{\alpha+\beta+2}} \{\alpha y - x(\beta + 1)\} \partial_{\ell,m}(x) dx \\
 &= \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1}}{(x+y)^{\alpha+\beta+2}} \{x(\beta + 1) - \alpha y\} \partial_{\ell,m}(x) dx
 \end{aligned}$$

the integrated term vanishing, since (1.10) holds. Since right hand side of (1.13) and (1.14) are the same, therefore the first equality of (1.6) follows.

By the first equation of (1.6)

$$\begin{aligned}
 y \frac{d}{dy} U_{\alpha,\beta,\ell,m}(y) &= -(\ell + m) [U_{\alpha,\beta,\ell,m}(y) - U_{\alpha,\beta,\ell-1,m}(y)] \\
 &= - \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1}}{(x+y)^{\alpha+\beta+1}} \left\{ -x \frac{d}{dx} \partial_{\ell,m}(x) \right\} dx \\
 &= \frac{\Gamma(\alpha + \beta + 1)}{\Gamma(\alpha)\Gamma(\beta + 1)} y^{\beta+1} \int_0^\infty \frac{x^{\alpha-1}}{(x+y)^{\alpha+\beta+1}} \tau_{\ell,m}(x) dx. \\
 &= B_{\alpha,\beta,\ell,m}(y)
 \end{aligned}$$

This proves the second equality of (1.6).

2.

For any number p , used an index (exponent), and such that $p > 1$, we write

$q = \frac{p}{p-1}$, so that p and q are conjugate indices in the sense of Holdr's inequality,

i.e. $\frac{1}{p} + \frac{1}{q} = 1$. Moreover, we define q to be if $p = 1$. We use $\{d, e, f, \dots\}$ to denote a positive constant depending only on the parameters d, e, f, \dots not necessarily on the two occurrences. Inequality of the form $L \leq c(d, e, f, \dots)R$ are to be interpreted as meaning "if the expression R is finite, then the expression L is finite and satisfies the inequality."

3. Known Results

The following results we due to Mishra and Singh [6].

Theorem A. If $\ell > n \geq 1, \alpha > 0, \beta > -1, m > -1$, then

$$s(x) \rightarrow (D, \alpha, \beta)(C, \ell - 1, m),$$

whenever $s(x) \rightarrow s(D, \alpha, \beta)(C, n - 1, m)$.

Theorem B. Let $\ell > n \geq 0, r \geq 0$ and $m > -1$. Suppose that $s(x)$ is summable (C, r, m) to s and that $\int_1^\infty \frac{\partial_{n,m}(x)}{x^{\beta+2}} dx$ converges, then $s(x)$ is summable $(D, \alpha, \beta)(C, \ell, m)$ to S .

4. DEFINITION OF STRONG SUMMABILITY BASED ON $(D, \alpha, \beta)(C, \ell, m)$ SUMMABILITY METHODS

Suppose that $\alpha > 0, \beta > -1, \ell \geq m \geq -1$ and $p \geq 1$, if

$$\int_0^x \left| U_{\alpha, \beta, \ell-1, m}(y) - s \right|^p dy = o(x) \text{ as } x \rightarrow \infty \quad (4.1)$$

Then we say that the function $s(x)$ summable $[(D, \alpha, \beta)(C, \ell, m)]_p$ or strongly summable $(D, \alpha, \beta)(C, \ell, m)$ with index p , to the sum s , if

$$\int_0^x \left| U_{\alpha, \beta, \ell-1, m}(y) \right|^p dy = O(x) \text{ as } x \rightarrow 0 \quad (4.2)$$

then we say that the function $s(x)$ is bounded $[(D, \alpha, \beta)(C, \ell, m)]_p$ or strongly bounded $(D, \alpha, \beta)(C, \ell, m)_p$ with index p .

When $p = 1$, $[(D, \alpha, \beta)(C, \ell, m)]_p$ and $[D, \alpha, \beta)(C, \ell, m)]$ denote the same method. The summability $(D, \alpha, \beta)(C, \ell - 1, m)$ may be regarded as the case $p = \infty$ of summability $[(D, \alpha, \beta)(C, \ell, m)]_p$ as is evident by the following equality

$$\lim_{p \rightarrow \infty} \left\{ x^{-1} \int_0^x \left| U_{\alpha, \beta, \ell, m}(y) - s \right|^p dy \right\}^{1/p} = \sup \left| U_{\alpha, \beta, \ell-1, m}(y) - s \right|^p.$$

We may also regard boundendness $(D, \alpha, \beta)(C, \ell - 1, m)$ as the case $p = \infty$ of boundedness $[(D, \alpha, \beta)(C, \ell - 1, m)]_p$.

5. SOME ELEMENTARY RESULTS

Theorem 5.1: If the function $s(x)$ is summable $[(D, \alpha, \beta)(C, \ell, m)]_p$ to the sum s , where $1 \leq p \leq \infty, \alpha > 0, \beta > -l, m > -1 \& \ell \geq 1$, then it is summable $(D, \alpha, \beta)(C, \ell, m)$ to the same sum.

In order to prove this Theorem, we require the following Lemma.

Lemma 5.1: Let $\ell > n \geq 1, \alpha > 0, \beta > -1, m > -1$, suppose that $I_{n-1, m}$ converges. Then for a given $x > 0$,

$$(I_{n-1, m} = \int_1^\infty \frac{\partial_{n-1, m}(x)}{x^{\beta+2}} dx)$$

$$U_{\alpha, \beta, \ell-1, m}(x) = \frac{\Gamma(\ell + m)}{\Gamma(\ell - n) \Gamma(n + m)} x^{-(\ell + m - 1)} \int_0^x y^{n+m-1} (x - y)^{\ell - n - 1} U_{\alpha, \beta, n-1, m}(y) dy \quad (5.1)$$

where (5.1) is to be taken as including the assertion that both sides exist. This Lemma has been proved by Mishra and Singh [6].

Proof of Theorem 5.1: We observed that, if the function $s(x)$ is summable $[D, \alpha, \beta)(C, \ell, m)]_p$ to the sum s , for any p such that $1 < p \leq \infty$, then it is summable $[D, \alpha, \beta)(C, \ell, m]$ to s . We may therefore suppose that $p = 1$, i.e.,

$$\int_0^x \left| U_{\alpha, \beta, \ell-1, m}(y) - s \right| dy = o(x)$$

Further, we may suppose that $s = 0$. Then replacing ℓ by $\ell + 1$ and n by $n + 1$ in

Lemma 5.1 we have

$$U_{\alpha, \beta, \ell, m}(x) \frac{\Gamma(\ell + m + 1)}{\Gamma(\ell - n) \Gamma(n + m + 1)} x^{-(\ell+m)} \int_0^x y^{n+m} (x - y)^{\ell - n - 1} U_{\alpha, \beta, n, m}(y) dy \quad (5.2)$$

Taking $n = \ell - 1$ in (5.2), we have

$$U_{\alpha, \beta, \ell, m}(x) = (\ell + m) x^{-(\ell+m)} \int_0^x y^{m+\ell-1} U_{\alpha, \beta, \ell-1, m}(y) dy$$

Now

$$|U_{\alpha, \beta, \ell, m}(x)| \leq (\ell + m) x^{-(\ell+m)} \int_0^x y^{m+\ell-1} |U_{\alpha, \beta, \ell-1, m}(y)| dy$$

Integrating by parts, we have

$$\begin{aligned} |U_{\alpha, \beta, \ell, m}(x)| &\leq (\ell + m) x^{-(\ell+m)} \left[\{o(x) y^{m+\ell-1}\}_0^x + (m + \ell - 1) \int_0^x o(x) y^{m+\ell-2} dy \right] \\ &= o(1). \end{aligned}$$

Hence $U_{\alpha, \beta, \ell, m}(x) = o(1)$.

This proves Theorem 5.1.

Theorem 5.2 : If the function $s(x)$ is summable $[(D, \alpha, \beta)(C, n, m)]_p$ to s , where $p > 1$, $\ell > n + \frac{1}{p}$, $n \geq 1$, then it is summable $(D, \alpha, \beta)(C, \ell - 1, m)$ to the sum

Proof : We may evidently suppose that $s = 0$. By Lemma 5.1, we have

$$U_{\alpha, \beta, \ell-1, m}(x) = \frac{\Gamma(\ell + m)}{\Gamma(\ell - n) \Gamma(n + m)} x^{-(\ell+m-1)} \int_0^x y^{n+m-1} (x - y)^{\ell - n - 1} U_{\alpha, \beta, n-1, m}(y) dy$$

Now

$$|U_{\alpha, \beta, \ell-1, m}(x)| \leq C x^{-(\ell+m-1)} \int_0^x y^{n+m-1} (x - y)^{\ell - n - 1} |U_{\alpha, \beta, n-1, m}(y)| dy$$

$$= Cx^{-(\ell+m-1)} \int_0^x \left\{ y^{\frac{n+m-1}{p}} |U_{\alpha, \beta, n-1, m}(y)| \right\} \{y^{\frac{n+m-1}{q}} (x-y)^{\ell-n-1}\} dy$$

Applying Höder's inequality with indices p and q , we have

$$\begin{aligned} |U_{\alpha, \beta, \ell-1, m}(x)| &\leq Cx^{-(\ell+m-1)} \left\{ \int_0^x y^{n+m-1} |U_{\alpha, \beta, n-1, m}(y)|^p dy \right\}^{1/p} \left\{ \int_0^x y^{n+m-1} (x-y)^{(\ell-n-1)q} dy \right\}^{1/q} \\ &= C \left\{ x^{-(\ell+m-1)} o(x) \right\}^{\frac{n+m-1}{p} + 1/p + \frac{n+m-1}{q} + (\ell-n-1) + 1/q} \\ &= o(x^0) = o(1) \end{aligned}$$

Provided that $\ell > n + 1/p$, $n \geq 1$.

Hence $U_{\alpha, \beta, \ell-1, m}(x) = o(1)$ as $x \rightarrow \infty$

This proves Theorem 5.2

Theorem 5.3 : Suppose that $\ell \geq 1$, $1 \leq p \leq q \leq \infty$ ($p < \infty$); $\ell > n + 1/p - 1/q$,

where $>$ maybe replaced by \leq in the last inequality in the case, $1 < p, q < \infty$;

or $1 = p, q = \infty$.

Then: (I) $[(D, \alpha, \beta)(C, n, m)]_p \Rightarrow [(D, \alpha, \beta)(C, \ell, m)]_q$.

(II) for any s

$$\begin{aligned} \sup_x \left\{ x^{-1} \int_0^x |U_{\alpha, \beta, \ell-1, m}(y) - s|^q dy \right\}^{1/q} &\leq C(p, q, \ell, m) \\ \sup_x \left\{ x^{-1} \int_0^x |U_{\alpha, \beta, n-1, m}(y) - s|^p dy \right\}^{1/p} \end{aligned} \tag{5.3}$$

The case $p = q = \infty$ is given by Theorem A.

For the proof of Theorem 5.3 we require few Lemmas.

Lemma 5.2 : Let $0 < \mu < 1$,

$$S(w) = \frac{1}{w} \int_0^w \phi(x) dx \quad (\phi(x) \geq 0)$$

and $s = \sup_{0 < w < \infty} S(w)$

Then $\int_0^w x^{-\mu} \phi(x) dx \leq c(\mu) s w^{1-\mu}$

If we take, $s(w) = o(1)$, as $w \rightarrow \infty$, then

$$\int_0^w x^{-\mu} \phi(x) dx = o(w)^{1-\mu}.$$

This Lemma is given by Mishra [4].

Lemma 5.3: If $f(x) \geq 0$ and $f(x) \in L^p(0, y)$, $p > 1$, and if

$$f_\alpha(x) = \frac{1}{\Gamma(\alpha)} \int_0^x (x-t)^{\alpha-1} f(t) dt,$$

where $0 < \alpha < \frac{1}{p}$ and $q = \frac{p}{1-p\alpha}$, then $f_\alpha(x) \in L^q(0, y)$

$$\text{and } \left\{ \int_0^y \{f_\alpha(x)\}^q dx \right\}^{1/q} \leq C \left\{ \int_0^y \{f(x)\}^p dx \right\}^{1/p},$$

where $C = C(p, \alpha) = C(p, q)$.

This Lemma has been proved by Hady, Littlewood and Polya [2].

We pass now to the proof of Theorem 5.3, we may evidently suppose that $s = 0$. Since Theorem 5.3 (I) is a corollary of Theorem 5.3 (II), it is sufficient to prove Theorem 5.3 (II) Let $\ell = n + \gamma$

Case (i) : $1 \leq p \leq q \leq \infty$

$$\text{and } \gamma > \frac{1}{p} - \frac{1}{q},$$

Case (ii) : $1 < p < q < \infty$

$$\text{and } \gamma = \frac{1}{p} - \frac{1}{q},$$

Case (iiia) : $1 < p < q = \infty$

$$\text{and } \gamma > \frac{1}{p} - \frac{1}{q}$$

Case (iiib) $1 = p < q = \infty$ and $\gamma \geq \frac{1}{p} - \frac{1}{q}$.

Proof of Case (i) Let S denote the superemum on the right of (5.3) suppose that

$$\frac{1}{r} = 1 - \frac{1}{p} + \frac{1}{q} \text{ and } 0 < \delta < 1, \text{ so that } r(\ell - n - 1) > -1 \quad r \geq 1.$$

By Lemma 5.1, We have

$$\begin{aligned} |U_{\alpha, \beta, \ell-1, m}(x)| &\leq Cx^{-(\ell+m-1)} \int_0^x y^{n+m-1} (x-y)^{\ell-n-1} |U_{\alpha, \beta, n-1, m}(y)| dy \\ &= Cx^{-(\ell+m-1)} \int_0^x \{(x-y)^{r(\ell-n-1)(1-\frac{1}{p})} y^{r(n+m-1+\delta)(1-\frac{1}{p})-\delta(1-\frac{1}{p})} \} \{y^{-\delta(\frac{1}{p}-\frac{1}{q})} |U_{\alpha, \beta, n-1, m}(y)|\}^{p(\frac{1}{p}-\frac{1}{q})} \\ &\quad \{x-y\}^{r(\ell-n-1)\frac{1}{q}} y^{r(n+m-1+\delta)\frac{1}{q}-\frac{\delta}{q}} |U_{\alpha, \beta, n-1, m}(y)|^{p/q} dy \end{aligned}$$

Applying Holder's inequality with indices $\frac{p}{p-1}$, $\frac{pq}{q-p}$ and q . we have

$$\begin{aligned} |U_{\alpha, \beta, \ell-1, m}(x)| &\leq Cx^{-(\ell+m-1)} \left\{ \int_0^x (x-y)^{r(\ell-n-1)} y^{r(n+m-1+\delta)-\delta} dy \right\}^{1-\frac{1}{p}} \left\{ \int_0^x y^{-\delta} |U_{\alpha, \beta, n-1, m}(y)|^p dy \right\}^{(\frac{1}{p}-\frac{1}{q})} \\ &\quad \left\{ \int_0^x (x-y)^{r(\ell-n-1)} y^{r(n+m-1+\delta)-\delta} |U_{\alpha, \beta, n-1, m}(y)|^p dy \right\}^{1/q} \\ &= Cx^{-(\ell+m-1)} x^{\{r(\ell-n-1)+r(n+m-1+\delta)-\delta+1\}(\frac{1}{p}-\frac{1}{q})} S^{p(\frac{1}{p}-\frac{1}{q})} \left\{ \int_0^x (x-y)^{r(\ell-n-1)} y^{r(n+m-1+\delta)-\delta} \right. \\ &\quad \left. |U_{\alpha, \beta, n-1, m}(y)|^p dy \right\}^{1/q} \quad (by \text{ lemma 5.2}) \end{aligned} \tag{5.4}$$

Hence

$$|U_{\alpha, \beta, \ell-1, m}(x)|^q \leq Cx^{-\{r(\ell+m-2+\delta)-\delta+1\}q} S^{q-p} \left\{ \int_0^x (x-y)^{r(\ell-n-1)} y^{r(n+m-1+\delta)-\delta} |U_{\alpha, \beta, n-1, m}(y)|^p dy \right\}^q$$

$$\int_0^w |U_{\alpha, \beta, \ell-1, m}(x)|^q dx \leq CS^{q-p} \int_0^w x^{-\{r(\ell+m-2+\delta)-\delta+1\}} \int_0^x (x-y)^{r(\ell-n-1)} y^{r(n+m-1+\delta)-\delta} |U_{\alpha, \beta, n-1, m}(y)|^p dy dx$$

Interchanging the order of integration, we have

$$\begin{aligned}
 \int_0^w |U_{\alpha, \beta, \ell-1, m}(x)|^q dx &\leq CS^{q-p} w^\delta \int_0^w (y^{r(n+m-1+\delta)-\delta} |U_{\alpha, \beta, n-1, m}(y)|^p)^p dy \int_y^w (x-y)^{r(\ell-n-1)} (x)^{-r(\ell+m-2+\delta-1)} dx \\
 &\leq CS^{q-p} w^\delta \int_0^w (y^{r(n+m-1+\delta)-\delta} |U_{\alpha, \beta, n-1, m}(y)|^p)^p dy \cdot y^{-r(n+m-1+\delta)} \\
 &\leq CS^{q-p} w^\delta \int_0^w y^{-\delta} |U_{\alpha, \beta, n-1, m}(y)|^p dy, \\
 &\leq CS^{q-p} w^\delta S^p w^{1-\delta}, \text{ by Lemma. 5.2} \\
 &\leq CS^q w
 \end{aligned}$$

Since

$$\begin{aligned}
 \int_x^w (y-u)^n y^{-n-\ell-1} dy &\leq \int_u^\infty (y-u)^n y^{-n-\ell-1} dy \\
 &= \frac{u^{-\ell} \Gamma(n+1) \Gamma(\ell)}{\Gamma(\ell+n+1)}, \quad (n > -1, u > 0, \ell > 0).
 \end{aligned}$$

Since $n+m-1 > -1$, we may choose δ such that $r(n+m-1+\delta) > 0$.

Thus by using Lemma 5.2, we have.

$$\sup_{w \geq 0} \{w^{-1} \int_0^w |U_{\alpha, \beta, \ell-1, m}(x)|^q dx\}^{1/q} \leq C(p, q, n, m) \sup_{w \geq 0} \{w^{-1} \int_0^w |U_{\alpha, \beta, n-1, m}(y)|^p dy\}^{1/p}$$

When $p = 1$, The first factor of (5.4) is to be omitted. Similarly when $q = p$, the second bracketed factor in (5.4) is to be omitted.

This proves Case (i)

Proof of Case (ii)

$$1 < p < q < \infty, \quad 0 < \gamma = \frac{1}{p} - \frac{1}{q} < \frac{1}{p}$$

By Lemma 5.1, we have

$$\begin{aligned}
 |U_{\alpha, \beta, \ell-1, m}(x)| &\leq C x^{-(n+\gamma+m-1)} \int_0^x y^{n+m-1} (x-y)^{\gamma-1} |U_{\alpha, \beta, n-1, m}(y)| dy \\
 &= C x^{-n-\gamma-m+1} \int_0^{x/2} y^{n+m-1} (x-y)^{\gamma-1} |U_{\alpha, \beta, n-1, m}(y)| dy + C x^{-n-\gamma-m+1} \int_{x/2}^x (x-y)^{\gamma-1} y^{n+m-1} |U_{\alpha, \beta, n-1, m}(y)| dy \\
 &= C(I_1 + I_2), \text{ say.}
 \end{aligned}$$

Hence, by Minkowski's inequality, it is sufficient to prove that

$$\sup_x \left\{ x^{-1} \int_0^x |I_1|^q dy \right\}^{1/q} \leq C \sup_x \left\{ x^{-1} \int_0^x |U_{\alpha, \beta, n-1, m}(y)|^p dy \right\}^{1/p} \quad (5.5)$$

and

$$\sup_x \left\{ x^{-1} \int_0^x |I_2|^q dy \right\}^{1/q} \leq C \sup_x \left\{ x^{-1} \int_0^x |U_{\alpha, \beta, n-1, m}(y)|^p dy \right\}^{1/p} \quad (5.6)$$

Proof of (5.5) :

Since $0 < \gamma < 1$, $(x-y)^{\gamma-1}$ increases in $(0, x/2)$, therefore we have

$$I_1 \leq C x^{-n-\gamma-m+1+\gamma-1} \int_0^{x/2} y^{n+m-1} |U_{\alpha, \beta, n-1, m}(y)| dy.$$

Proceeding as in the case (i), we see that (5.5) is established.

Proof of (5.6) :

$$I_2 \leq C \int_{x/2}^x (x-y)^{\gamma-1} y^{-\gamma} |U_{\alpha, \beta, n-1, m}(y)| dy.$$

Applying Lemma 5.3, we have

$$\left\{ \int_0^x |I_2|^q dy \right\}^{1/q} \leq C \left\{ \int_0^x y^{-(1-\frac{p}{q})} |U_{\alpha, \beta, n-1, m}(y)|^p dy \right\}^{1/p}$$

Since $1-p/q < 1$, therefore we can apply Lemma 5.2. Thus.

$$\sup_x \left\{ x^{-1} \int_0^x |U_{\alpha, \beta, \ell-1, m}(y)|^q dy \right\}^{1/q} \leq C(p, q, n) \sup_x \left\{ x^{-1} \int_0^x |U_{\alpha, \beta, n-1, m}(y)|^p dy \right\}^{1/p}.$$

This proves the Case (ii)

Proof of Case (iiia) is given by Theorem 5.2 Case (iiib) is given by Theorem 5.1. Thus, we have completely proved Theorem 5.3.

Theorem 5.4: The necessary and sufficient conditions for the function $s(x)$ to be summable $[(D, \alpha, \beta)](C, l, m)]_p$ to s are that it be summable $[(D, \alpha, \beta)](C, l, m)]_p$ to s and

$$\int_0^x \left| y \frac{d}{dy} U_{\alpha, \beta, \ell, m}(y) \right|^p dy = o(x) \text{ as } x \rightarrow \infty. \quad (\ell > 1, m > -1).$$

Proof. This result is analogous to Hyslop's Theorem 3 [1] on strong Cesaro summability. Since summability $[(D, \alpha, \beta)](C, l, m)]_p$ implies summability $[(D, \alpha, \beta)](C, l, m)]$ it follows from (1.7) and Minkowski's inequality, that since $U_{\alpha, \beta, \ell, m}(y) - s = o(1)$ as $y \rightarrow \infty$.

$$\begin{aligned} \left\{ \int_0^x \left| U_{\alpha, \beta, \ell-1}(y) - s \right|^p dy \right\}^{1/p} - \left\{ \int_0^x \left| U_{\alpha, \beta, \ell, m}(y) - s \right|^p dy \right\}^{1/p} &\leq \frac{1}{|\ell + m|} \left\{ \int_0^x \left| y \frac{d}{dy} U_{\alpha, \beta, \ell, m}(y) \right|^p dy \right\}^{1/p} \quad (5.7) \\ &\leq \left\{ \int_0^x \left| U_{\alpha, \beta, \ell-1, m}(y) - s \right|^p dy \right\}^{1/p} + \left\{ \int_0^x \left| U_{\alpha, \beta, \ell, m}(y) - s \right|^p dy \right\}^{1/p} \\ &= o(x)^{1/p} + o(x)^{1/p} \\ &= o(x)^{1/p} \end{aligned}$$

This implies that $\int_0^x \left| y \frac{d}{dy} U_{\alpha, \beta, \ell, m}(y) \right|^p dy = o(x)$ as $x \rightarrow \infty$

Conversely suppose that function $s(x)$ is summable $[(D, \alpha, \beta)](C, l, m)]$ to s and

$$\int_0^x \left| y \frac{d}{dy} U_{\alpha, \beta, l, m}(y) \right|^p dy = d(x) \text{ as } x \rightarrow \infty$$

Again using (5.7), we have

$$\int_0^x \left| U_{\alpha, \beta, \ell-1, m}(y) - s \right|^p dy = o(x) \text{ as } x \rightarrow \infty$$

This completes the proof of Theorem 5.4.

6. RELATIVE STRENGTH BETWEEN $[D, \alpha, \beta](C, \ell, m)_p$ AND $[C, n, m]_p$

Theorem 6.1: Suppose that $1 \leq p \leq \infty$ and $n \geq 0$. If the function $s(x)$ is summable $[C, n, m]_p$ to the sum s , it is summable $[D, \alpha, \beta](C, \ell, m)_p$ to the sum s for every $t \geq 1$, $\ell \geq 1$. Wherever $(D, \alpha, \beta)(C, \ell-1, m)$ method is applicable.

The proof of this theorem follows immediately from B and the chain of implications $[C, n, m]_p \Rightarrow (C, n, m) \Rightarrow (D, \alpha, \beta)(C, \ell-1, m) \Rightarrow [(D, \alpha, \beta)(C, \ell, m)]_p$ for every $t \geq 1$,

Theorem 6.2 Suppose $1 \leq p \leq \infty$ and $\ell \geq 1, m > -1, \ell > \delta > 0$. If the function $s(x)$ is bounded $[(D, \alpha, \beta)(C, \ell-\delta, m)]_p$ and summable $[D, \alpha, \beta, \ell-\delta+1, m]_p$ to the sum s , it is summable $[D, \alpha, \beta, \ell, m]_p$ to s .

Proof. We may suppose that $s = 0$, $0 < \delta < 1/p$. By Lemma 5.1 replacing n by $\ell - \delta$ we have

$$\begin{aligned} U_{\alpha, \beta, \ell-1, m}(x) &= Cx^{-(\ell+m-1)} \int_0^x (x-y)^{\ell-\ell+\delta-1} y^{\ell-\delta+m-1} U_{\alpha, \beta, \ell-\delta-1, m}(y) dy \\ &= Cx^{-(\ell+m-1)} \int_0^{x(1-\rho)} (x-y)^{\delta-1} y^{\ell-\delta+m-1} U_{\alpha, \beta, \ell-\delta-1, m}(y) dy + Cx^{-(\ell+m-1)} \int_{x(1-\rho)}^x (x-y)^{\delta-1} y^{\ell-\delta+m-1} U_{\alpha, \beta, \ell-\delta-1, m}(y) dy. \\ &= I_1(x) + I_2(x), \text{ say,} \end{aligned}$$

ρ being a fixed number such that $0 < \rho < 1/2$. Let now

$$S = \sup_x \left\{ x^{-1} \int_0^x \left| U_{\alpha, \beta, \ell-\delta-1, m}(y) \right|^p dy \right\}^{1/p}. \quad (6.1)$$

$$\text{We have } \left| I_2(x) \right|^p \leq C \left\{ x^{-(\ell+m-1)} \int_{x(1-\rho)}^x (x-y)^{\delta-1} y^{\ell-\delta+m-1} \left| U_{\alpha, \beta, \ell-\delta-1, m}(y) \right| dy \right\}^p$$

$$\leq C \left\{ \int_{x(1-\rho)}^x (x-y)^{\delta-1} y^{-\delta} \left| U_{\alpha, \beta, \ell-\delta-1, m}(y) \right| dy \right\}^p$$

or

$$\left| I_2(x) \right|^p \leq C \left\{ \int_{x(1-\rho)}^x (x-y)^{\frac{\delta-1}{p}} y^{-\delta} (x-y)^{\frac{\delta-1}{q}} \left| U_{\alpha, \beta, \ell-1, m}(y) \right| dy \right\}^p$$

Applying Holder's inequality with indices p and q , we have

$$\begin{aligned} \left| I_2(x) \right|^p &\leq C \left\{ \int_{x(1-\rho)}^x (x-y)^{\delta-1} y^{-\delta p} \left| U_{\alpha, \beta, \ell-1, m}(y) \right|^p dy \right\} \left\{ \int_{x(1-\rho)}^x (x-y)^{\delta-1} dy \right\}^{p/q} \\ &= C(\ell, m, \delta, p) x^{\delta(p-1)} \rho^{\delta(p-1)} - \int_0^t x^{\delta(p-1)} \int_{x(1-\rho)}^x (x-y)^{\delta-1} y^{-\delta p} \left| U_{\alpha, \beta, \ell-1, m}(y) \right|^p dy. \end{aligned}$$

Hence

$$\begin{aligned} \int_0^t \left| I_2(x) \right|^p dx &\leq C \rho^{\delta(p-1)} - \int_0^t \left\{ x^{\delta(p-1)} \int_{x(1-\rho)}^x (x-y)^{\delta-1} y^{-\delta p} \left| U_{\alpha, \beta, \ell-\delta-1, m}(y) \right|^p dy \right\} dx \\ &\leq C t^{\delta(p-1)} \rho^{\delta(p-1)} \int_0^t dx \int_{x(1-\rho)}^x (x-y)^{\delta-1} y^{-\delta p} \left| U_{\alpha, \beta, \ell-\delta-1, m}(y) \right|^p dy. \end{aligned}$$

Interchanging the order of integration, we have

$$\begin{aligned} \int_0^t \left| I_2(x) \right|^p dx &\leq C t^{\delta(p-1)} \rho^{\delta(p-1)} \int_0^t y^{-\delta p} \left| U_{\alpha, \beta, \ell-\delta-1, m}(y) \right|^p dy \int_y^{y/1-\rho} (x-y)^{\delta-1} dx \\ &= C t^{\delta p} \rho^{\delta p} (1-\rho)^\delta \int_0^t y^{-\delta p} \left| U_{\alpha, \beta, \ell-\delta-1, m}(y) \right|^p dy \end{aligned}$$

Applying Lemma 5.2 we have

$$\int_0^t \left| I_2(x) \right|^p dx \leq C S^p \rho^{\delta p} (1-\rho)^{-\delta} t,$$

(since $\delta p < 1$), the constant being independent of ρ . Replacing ℓ by $n+1$ in Lemma 5.1, we have

$$U_{\alpha, \beta, n, m}(x) = C x^{-(n+m)} \int_0^x y^{n+m-1} (x-y)^{n+1-n-1} U_{\alpha, \beta, n-1, m}(y) dy$$

or

$$U_{\alpha, \beta, n, m}(x) = C x^{-(n+m)} \int_0^x y^{n+m-1} U_{\alpha, \beta, n-1, m}(y) dy$$

we have

$$I_1(x) = C x^{-(\ell+m-1)} \int_0^{x(1-\rho)} (x-y)^{\delta-1} y^{\ell-\delta+m-1} U_{\alpha, \beta, \ell-\delta, m}(y) dy$$

Integrating by parts

$$\begin{aligned} I_1(x) &= C x^{-(\ell+m-1)} [(x-y)^{\delta-1} y^{\ell-\delta+m} U_{\alpha, \beta, \ell-\delta, m}(y)]_0^{x(1-\rho)} + C x^{-(\ell+m-1)} \int_0^{x(1-\rho)} (\delta-1)(x-y)^{\delta-2} y^{\ell-\delta+m} U_{\alpha, \beta, \ell-\delta, m}(y) dy \\ &= C x^{-(\ell+m-1)} [(\rho x)^{\delta-1} x^{\ell-\delta+m} (1-\rho)^{\ell-\delta+m} U_{\alpha, \beta, \ell-\delta, m}\{x(1-\rho)\}] + C x^{-(\ell+m-1)} (\delta-1) \int_0^{x(1-\rho)} (x-y)^{\delta-2} y^{\ell-\delta+m} U_{\alpha, \beta, \ell-\delta, m}(y) dy \\ &= C(1-\rho)^{\ell-\delta+m} \rho^{\delta-1} U_{\alpha, \beta, \ell-\delta, m}\{x(1-\rho)\} + C x^{-(\ell+m-1)} \int_0^{x(1-\rho)} (x-y)^{\delta-2} y^{\ell-\delta+m} U_{\alpha, \beta, \ell-\delta, m}(y) dy \\ &= I_{1,1}(x) + I_{1,2}(x), \text{(say).} \end{aligned}$$

$$I_{1,1}(x) = C \rho^{\delta-1} (1-\rho)^{\ell-\delta+m} U_{\alpha, \beta, \ell-\delta, m}\{x(1-\rho)\}$$

$$|I_{1,1}(x)|^p \leq C \rho^{\delta p - p} (1-\rho)^{p(\ell-\delta+m)} |U_{\alpha, \beta, \ell-\delta, m}\{x(1-\rho)\}|^p$$

Integration both sides, we get

$$\begin{aligned} \int_0^t |I_{1,1}(x)|^p dx &\leq C(\ell, m, \delta, \rho, p) \int_0^t |U_{\alpha, \beta, \ell-\delta, m}\{x(I-\rho)\}|^p dx \\ &= o(t). \end{aligned} \tag{6.2}$$

$$|I_{1,2}(x)|^p = C \left| x^{-(\ell+m-1)} \int_0^{x(1-\rho)} (x-y)^{\delta-2} y^{\ell-\delta+m} U_{\alpha, \beta, \ell-\delta, m}(y) dy \right|^p$$

if $p > 1$ we have, by Holder's inequality with indices p and q ,

$$\begin{aligned}
\left| I_{1,2}(x) \right|^p &\leq C x^{-(\ell+m-1)p} \left\{ \int_0^{x(1-\rho)} \left| U_{\alpha, \beta, \ell-\delta, m}(y) \right|^p y^{(\ell-\delta+m)p} dy \right\} \left\{ \int_0^{x(1-\rho)} \{(x-y)^{\delta-2}\}^q dy \right\}^{p/q} \\
\left| I_{1,2}(x) \right|^p &\leq C x^{-(\ell+m-1)p} \left\{ \int_0^{x(1-\rho)} \left| U_{\alpha, \beta, \ell-\delta, m}(y) \right|^p y^{(\ell-\delta+m)p} dy \left[\frac{-(x-y)^{(\delta-2)q+1}}{(\delta-2)q+1} \right]_0^{x(1-\rho)} \right\}^{p/q} \\
&\leq C x^{-(\ell+m-1)p} \int_0^{x(1-\rho)} \left| U_{\alpha, \beta, \ell-\delta, m}(y) \right|^p y^{(\ell-\delta+m)p} dy \left\{ \frac{[-(\rho x)^{(\delta-2)q+1} - x^{(\delta-2)q+1}]^{p/q}}{[(\delta-2)q+1]^{p/q}} \right\} \\
&\leq C x^{-(\ell+m-1)p} \int_0^{x(1-\rho)} \left| U_{\alpha, \beta, \ell-\delta, m}(y) \right|^p y^{(\ell-\delta+m)p} dy \left\{ \frac{x^{[(\delta-2)q+1]p/q} [\rho^{(\delta-2)q+1} - 1]^{p/q}}{[(\delta-2)q+1]^{p/q}} \right\} \\
&\leq C x^{-(\ell+m-1)p} \int_0^{x(1-\rho)} \left| U_{\alpha, \beta, \ell-\delta, m}(y) \right|^p y^{(\ell-\delta+m)p} dy x^{\delta p - p - 1} \quad (\text{since } 1/p + 1/q = 1) \\
&\leq C x^{-(\ell+m-1)p + \delta p - p - 1} \int_0^{x(1-\rho)} \left| U_{\alpha, \beta, \ell-\delta, m}(y) \right|^p y^{(\ell-\delta+m)p} dy \\
&\leq C x^{-\ell p - m p + \delta p - p - 1} \int_0^{x(1-\rho)} \left| U_{\alpha, \beta, \ell-\delta, m}(y) \right|^p y^{(\ell-\delta+m)p} dy \\
&\leq C x^{-1} \int_0^{x(1-\rho)} \left| U_{\alpha, \beta, \ell-\delta, m}(y) \right|^p dy \\
&= O(x^{-1}) o(x) = o(1).
\end{aligned}$$

If $p = 1$, we have

$$\begin{aligned}
\left| I_{1,2}(x) \right| &\leq O(x^{-\ell-m+\delta-1}) \int_0^{x(1-\rho)} \left| U_{\alpha, \beta, \ell-\delta, m}(y) \right| y^{\ell-s+m} dy \\
&\leq O(x^{-\ell-m+\delta-1}) o(x^{\ell-\delta+m+1}) \\
&= o(1),
\end{aligned}$$

and so in either case $\int_0^t \left| I_{1,2}(x) \right|^p dx \leq C \int_0^t o(1) dx = o(t)$ (6.3)

It follows by Minkowski's inequality that

$$\begin{aligned} & \limsup \left\{ t^{-1} \int_0^t \left| U_{\alpha, \beta, \ell-1, m}(x) \right|^p dx \right\}^{1/p} \\ & \leq SC(\ell, m, \delta, p) \rho^\delta (1-\rho)^{-\delta/p} + o(1) + o(1) \\ & \leq SC(\ell, m, \delta, p) \rho^\delta (1-\rho)^{-\delta/p} \end{aligned} \quad (6.4)$$

Making $\rho \rightarrow 0$, it follows now that limit sup on the left of (6.4) is 0, and this completes the proof of Theorem 6.2.

Theorem 6.3: Suppose that $1 < p \leq \infty, \alpha > 0, \beta \geq -1, m > -1, \ell > 1$ and (1.5) holds. If the function $s(x)$ is summable $[D, \alpha, \beta](C, \ell, m)_p$ to the sum s , then the function $xa(x)$ is summable $[(D, \alpha, \beta)(C, \ell+1, m)]_p$ to the sum 0.

Theorem 6.4: Suppose that $1 \leq p \leq \infty, \alpha > 0, \beta \geq -1$, and (1.5) holds. If the function $s(x)$ is summable $[(D, \alpha, \beta)(C, \ell+1, m)]_p$ to s and $xa(x)$ is summable $[(D, \alpha, \beta)(C, \ell+1, m)]_p$ to the sum 0, then the function $s(x)$ is summable $[(D, \alpha, \beta)(C, \ell, m)]_p$.

Theorem 6.5: Suppose that $1 \leq p \leq \infty, \alpha > 0, \beta \geq -1, \delta > 0$ and (1.5) holds. If the function $s(x)$ is summable $[(D, \alpha, \beta)(C, \ell-\delta, m)]_p$ to s and the function $xa(x)$ is bounded $[(D, \alpha, \beta)(C, \ell-\delta, m)]_p$ then the function $s(x)$ is summable $[(D, \alpha, \beta)(C, 1-\delta, m)]_p$ to the sum s .

Proof of Theorem 6.3 follows immediately Theorem 5.4 while Theorem 5.4 follows directly from (1.6) and Minkowski's inequality.

To prove Theorem 6.5, we first note that if the function $s(x)$ is summable $[(D, \alpha, \beta)(C, \ell-\delta, m)]_p$ then the function $xa(x)$ is summable $[(D, \alpha, \beta)(C, \ell-\delta+1, m)]_p$. If also the function $xa(x)$ is bounded $[(D, \alpha, \beta)(C, \ell-\delta, m)]_p$, then it is summable $[(D, \alpha, \beta)(C, \ell, m)]_p$ to 0, and since summability $[(D, \alpha, \beta)(C, \ell-\delta, m)]_p$ implies summability $[(D, \alpha, \beta)(C, \ell, m)]_p$, Theorem 6.4 completes the proof of Theorem 6.5.

REFERENCES

1. J.M. Hyslop: Note on the strong summability of series. Proc. Glasgow Math. Assoc. 1, (1951/53), 16-20.
2. G.H. Hardy, J.E. Littlewood, and G. Polya : Inequalities. (Cambridge, 1959), P. 290.
3. B. Kwee : On Generalized translated quasi-cesaro summability. Pacific Journal of Mathematics 36, (No. 3) (1971) 731-740.
4. B.P. Mishra. : Strong summability of infinite series on a scale of Abel type summability methods. Proc. Cambridge Philos. Soc. 63 (1967) 119-127
5. B.P. Mishra & A.P. Srivastava: some remarks on absolute summability of function based on (C, α, β) summability methods, Nat. Acad. of Mathematics 13 (1999).
6. B.P. Mishra. & S.K. Singh: On product summability methods $((D, \alpha, \beta)(C, \ell, m)$ of function communicated for publication,
7. Singh, B.N. : Some remarks on summability methods function, Ph. D. Theses, Gorakhpur University (1992).

EFFECT OF POLARIZABILITY ON NUCLEATION PHENOMENON DURING ICE GLACIATION DUE TO EXTERNAL ELECTRIC FIELD

N.SINGH*, DEVENDRA SINGH, VIKAS MISHRA** AND PRAKASH MISHRA****

(Received 14-08-2004)

ABSTRACT

The polarizability of water vapour molecules in presence of external electric field plays an important role in the nucleation rate of water vapour condensation and ice glaciation. The polarizability decreases with increase in the temperature and hence there is decrease in Gibb's free energy and hence nucleation rate is decreased. Thus polarizability controls the ice glaciation in clouds.

Keywords and phrases: Polarization. Nucleation rate ,Gibb's free energy, Critical nuclei.

INTRODUCTION

The polarizability plays a dominant role in inducing the electric dipole on water vapour molecules. The contribution to polarization of water vapour molecules coming from the electric field generated by the droplet dipole, which increases the interaction between droplet and water vapour molecules. This would enhance the effect described by Murino [18], but the interaction due to this additional factor would not be steady and hence would produce only transient effect.

Ice nucleation is a fundamental cloud process. The process involved in ice formation is understood only in the case of homogeneous freezing of supercooled droplets. In contrast, ice formation caused by heterogeneous freezing nuclei at low temperatures are poorly understood. Kärcher and Lohmann [12] considered theoretically the nucleation and initial growth of ice crystals in cirrus clouds at

*Department of Physics, Nehru college, Chhatarpur-200021, U.P.

**Department of Physics, D.B.S. College, Kanpur-208006, U.P.

low temperatures ($< 235K$) prevailing in the upper troposphere and in the tropopause region.

Ice crystals are produced by electric fields in cloud chambers [8] caused by fragmentation of dendritic crystals. The nucleation of ice in supercooled water by electric fields has been demonstrated. Nucleation in the thin liquid filaments, which form during disruption of drops or during movement of drops on surfaces was suggested by Loeb [13] and by Abbas and Latham [1] as a possible mechanism.

Nucleation as a result of distortion of drops by electric forces was reported [1,20,21,28]. When charged, aerosols and other surfaces coming in contact with drops were found to cause nucleation more readily than if no electric charge were present [9,21,23]. The ice formation in the atmosphere is a heterogeneous process initiated by ice forming nuclei [15].

Dufour [7] suggested that freezing could be initiated in supercooled water by means of an electric field. Unfortunately he gave only an insufficient description of his experiment so that no firm conclusions can be drawn. Dufour [7], Rau [22], Salt [25], Gabarashvili and Gliki [9] proposed that electro-freezing of supercooled water is due to dielectric polarization of water in high electric fields.

Collins [4] inferred that the relaxation time is independent of the free energy of formation of the nucleus, but it varies as the square of the radius of the critical nucleus. Bartlett et al. [3], as well as Maybank and Barthakur [16] reported the tendency of generation of a few ice splinters during ice crystal growth in strong electric fields. Doolittle and Vali [6] pointed out the effect of an electric field on samples of water at different time temperature cycles.

The formulation of relaxation time required for the attainment of the steady state concentrations of the embryos of the critical size has been discussed [4,11,19,29]. The scavenging of aerosol particles by hydro-meteors under external electric field has been discussed by Wang [30]. Sharma et al. [26] reported that the small values of electric field is equivalent to very high supersaturation ratio to get a nucleus of given size under similar conditions of temperature.

Singh et al. [27] have shown that the resultant electric effect on a droplet

due to an external electric field and the field induced due to the central dipole, the role of nucleation in water vapour condensation and ice glaciation is about hundred times more near breakdown for dry air as compared to that in absence of electric field. Thunderstorm electric field promote droplet growth.

Electric field near triggered lightning channels have been measured with Pockels sensors. Mikki et al. [17] at the International Centre for Lightning Research and Testing at Camp Blanding, Florida. The vertical electric field pulse peaks are in the range from 176 kV/m to 1.5 MV/m (the median is 577 kV/m) and horizontal electric field pulse peaks are in the range from 495 kV/m to 1.2 MV/m (the median is 821 kV/m). The vertical electric field measured very close to the lightning channel tends to increase with an increase in the previous no current interval.

The atmospheric electric field and the electrical conductivity varies from place to place depending upon certain environmental factors. Deshpande and Kamra [5] gave the surface measurements of the atmospheric electric potential gradient and conductivity at Indian station, Maitri ($70^{\circ} 45' 52''$ S, $11^{\circ} 44' 3''$ E, 117 m above mean sea level), Antarctica, from January 10 to February 24, 1997. The diurnal variation of potential gradient averaged for 20 fair weather days is single periodic with maximum at 1300 UT and another secondary but distinct maximum at 1900 UT and a minimum at 0100 UT. The mean value or the potential gradient is 83 V/m and the maximum and minimum values are 1.55 and 0.73 times the mean value, respectively. The total conductivity is 2.1×10^{-14} S/m and does not show any significant diurnal variation. The diurnal variation of the potential gradient observed at Maitri is similar to the one observed a few days later in the southern Indian Ocean and to the one reported by Kamra et al. [10] with a few hours difference in phase in the northern Indian Ocean.

Rust and Trapp [24] presented the first known vertical profiles of electric field in six winter nimbo stratus clouds in the U.S.A. The maximum magnitude of the vertical component of electric field in the profiles ranged from 1 to 12 kV/m, the maximum horizontal component ranged from 0.2 to 28 kV/m.

CC-0. In Public Domain. Gurukul Kangri Collection, Haridwar
Marshall and Stolzenburg [14] used a one-dimensional model to calculate

the electrostatic energy of several horizontally extensive, electrified clouds. Model estimates of the model calculation are initiated with known charge distributions. Small reductions in the charge density of a region often can provide a relatively large amount of energy to derive a flash. Model calculations suggest that positive cloud-to-ground flashes that produce Q-bursts have energies of about 1×10^{10} J. Total electrostatic energy stored in two stratiform clouds of mesoscale convective systems were 5×10^{11} J and 2×10^{12} J. These energies are sufficient to support hundreds or thousands of typical lightning flashes, but only 10-100 of the energetic positive cloud-to-ground flashes with associated Q-bursts.

To authors knowledge the role of effective polarizability has not been considered affecting nucleation rate of ice particles. In present study, we have evaluated the effect of polarizability on nucleation effective polarizability. We estimated nucleation rate of ice particles in external electric field using modified value of polarizability at different temperatures. Due to the effective polarizability α_{eff} , the Gibb's free energy, nucleation rate and radius of nucleus of ice are found to increase.

THEORETICAL CONSIDERATION

The effective or resultant polarizability may be written as

$$\alpha_{eff} = \alpha + \frac{p_0^2}{3kT} \quad (1)$$

Thus, effective polarizability varies inversely as absolute temperature of medium. In an external electric field, the moment induced on a droplet is

$$\bar{M} = \bar{E}r_i^3 \quad (2)$$

where, r_i is the radius of ice embryo (assumed spherical) and \bar{E} , the electric field.

The moment induced on water vapour molecule, present in the cloud is

$$\vec{M}' = \alpha \vec{E} \quad (3)$$

where, α is the polarizability. For α we assume the value $5 \times 10^{-23} \text{ cm}^3$, observing that the contribution from the permanent electric dipole to polarizability is

$$\frac{p_0^2}{3kT} = 2.6 \times 10^{-23} \text{ cm}^3$$

where, p is the permanent dipole moment; k , the Boltzmann constant and T , the absolute temperature.

HOMOGENEOUS NUCLEATION

In the classical models of drop condensation and ice nucleation the Gibbs function has been taken into consideration. However, Abraham[2] has shown that the Helmholtz free energy is the proper thermodynamic potential and Gibbs function is only its approximation.

The Gibb's free energy for formation of a ice embryo is given by

$$\Delta G_i = -\frac{4}{3}\pi r_i^3 \Delta G_v + 4\pi r_i^2 \sigma_{i/v} \quad (4)$$

$$\text{where } \Delta G_v = \frac{\rho_i RT \ln S_{v,i}}{M_w}$$

R , the universal Gas constant; $S_{v,i}$, the supersaturation ratio of water vapour over ice surface; M_w , the molecular weight of water.

Critical radius of ice nucleus is given by

$$\frac{\partial \Delta G_i}{\partial r_i} = 0$$

The radius of critical ice nucleus is

$$r_i^* = 2\sigma_{i/v} M_w / \rho_i RT \ln S_{v,i} \quad (5)$$

The number ice molecules in a critical nucleus is given by

$$n_i^* = 4\pi r_i^{*3} \rho_i N / 3M_w \quad (6)$$

where, ρ is density of ice; N , the Avogadro number.

The energy of formation of a critical nucleus is

$$\Delta G_i^* = 4\pi r_i^{*2} \sigma_{i/v} / 3 \quad (7)$$

The nucleation rate of ice nuclei is

$$J_i^* = \text{a constt.} \exp[-\Delta G_i^* / kT] \quad (8)$$

IN PRESENCE OF EXTERNAL ELECTRIC FIELD

Water is strongly polarizable having dipole moment 1.83×10^{-18} esu cm. In an external electric field, the water droplet (or embryo) rapidly increases in size. The expression for the rate of growth of radius of ice embryo has been derived as [27].

$$\frac{dr_i}{dt} = \frac{\rho_v}{\rho_i} \left[9\alpha_{eff} \lambda E^2 / m_w r_i \right]^{1/2} \quad (9)$$

where, λ is the mean free path; E , the external electric field; m_w , the mass of water vapour molecule; ρ_v , the density of water vapour and r_i the radius of ice and embryo in the electric field. Integrating above equation within limits $r_i = 0$ to r_i^* (critical radius of ice nucleus) and $t = 0$ to $t = \tau$ (relaxation time)

$$r_i^* = \left[3\rho_v \tau \left(9\alpha_{eff} \lambda E^2 / m_w \right)^{1/2} / 2\rho_i \right]^{2/3} \quad (10)$$

Analogically, in the external field eqn. (3.32) and (3.33) for the energy of formation and the nucleation rate of ice nucleus, becomes

$$\Delta G_i^* = \frac{4\pi}{3} r_i^* \sigma_{i/v} \quad (11)$$

and

$$J_i = \text{a constant. } \exp\left(-\frac{\Delta G_i^*}{kT}\right) \quad (12)$$

RESULTS

The effective polarizability α_{eff} of water varies nearly inversely as the absolute temperature. The variation of α_{eff} with temperature has been shown in table 1.

Table 1. : Calculated values of α_{eff} of water molecules with the absolute temperature using $\alpha = 5 \times 10^{-23} \text{ cm}^3$.

T(K)	$\alpha_{eff} (x 10^{-23}) \text{ cm}^3$
263	8.006
273	7.896
283	7.794
293	7.698
303	7.609

Using eqn. (9), (10) and (11) the critical radius of the ice nucleus r_i^* , Gibb's free energy of formation ΔG_i^* and natural logarithm of nucleation rate $\ln J_i^*$ as the function of external electric field at relaxation time 2×10^{-6} sec at temperatures 263, 273 and 283 K have been shown in Table 2.

Calculated value of critical radius of ice, r_i^* , Gibb's free energy of formation of critical ice nucleus, ΔG_i^* , and natural logarithm of nucleation rate of ice nuclei $\ln J_i^*$ as the function of the relaxation time at constant electric field 10 esu at temperatures 263, 273 and 283 K are given in Table 3.

Table 2. : Calculated values of r_i^* , ΔG_i^* and $\ln J_i^*$ as the function of external electric field at $\tau = 2 \times 10^{-6}$ s and temperatures 263, 173 and 283 K.

E (esu)	$r_i^* (\times 10^{-9})$ cm	$\Delta G_i^* (\times 10^{-14})$ erg	$\ln J_i^*$
T = 263 K			
5	18.59	14.48	3.99
10	29.50	36.48	10.05
15	38.66	62.63	17.26
T = 273 K			
5	18.51	14.35	3.81
10	29.38	36.16	9.60
15	38.49	62.09	16.48
T = 283 K			
5	18.43	14.23	3.64
10	29.25	35.85	9.18
15	38.33	61.55	15.76

Table 3. : Calculated values of r_i^* , ΔG_i^* and $\ln J_i^*$ as the function of relaxation time at E = 10 esu at temperatures 263, 173 and 283 K.

T ($\times 10^{-9}$) sec	$r_i^* (\times 10^{-9})$ cm	$\Delta G_i^* (\times 10^{-14})$ erg	$\ln J_i^*$
T = 263 K			
10	86.27	30.92	82.94
20	136.94	78.58	216.52
30	179.44	134.93	371.78
T = 273 K			
10	85.90	30.65	82.06
20	136.35	76.23	206.79
30	178.67	133.77	355.08
T = 283 K			
10	85.52	30.10	78.48
20	135.76	73.91	197.75
30	177.89	132.61	339.56

DISCUSSION

The calculations in Table 1, 2 and 3 have been made using the following constants:

ρ_v	=	10^{-5} at 10^0 gm cm^{-3}
α	=	5.0×10^{-23} cm^3
p	=	1.81×10^{-18} esu
k	=	1.38×10^{-16} erg K^{-1}
λ	=	10^{-5} cm
m_w	=	3.0×10^{-23} gm
M_w	=	18
ρ_i	=	0.917 gm cm^{-3}
$\rho_{i/v}$	=	100 dynes cm^{-1}
R	=	8.317×10^7 erg K^{-1} mole $^{-1}$
N_A	=	6.025×10^{23}

The effective polarizability of water varies approximately inversely as absolute temperature (Table 1). From Table 2 it is clear that radius decreases with increase in temperature. Also Gibb's free energy decreases with increase in temperature.

From Table 2, we see that at the given temperature, the value of r_i^* , ΔG_i^* and $\ln J_i^*$ are found to increase with increase in applied external electric field. But, at a fixed external electric field, the values of r_i^* , ΔG_i^* and $\ln J_i^*$ decrease with increase in temperature.

For example, at electric field 10 esu, the radii are 29.50×10^{-9} , 29.38×10^{-9} and 29.25×10^{-9} cm corresponding to the temperatures 263, 273 and 283 K, respectively. Similarly, at the same electric field, the critical Gibb's free energy for formation of ice nucleus are 36.48×10^{-14} , 36.16×10^{-14} , and 35.85×10^{-14} erg at temperatures 263, 273 and 283 K, respectively. Similar variations are evident for $\ln J_i^*$.

Table 3 represents the calculated values of r_i^* , ΔG_i^* and $\ln J_i^*$ as the function of relaxation time at electric field 10 esu for temperature 263, 273 and 283 K. These values have been found to increase with increase in relaxation time for given temperature. But at a given relaxation time, these values decrease with increase in temperature.

From above discussion we may conclude that the nucleation parameters, critical radius, Gibb's free energy are highly decreased, while nucleation rate is increased with increase in temperature. Thus, the nuclei of smaller size requiring less Gibb's free energy are produced in large numbers i.e. nucleation rate is enhanced.

CONCLUSION

From above study, it may be concluded that the nucleation process is accelerated in presence of external electric field as compared with the homogeneous nucleation process. In clouds and thunderstorms, electric fields are generated due to lightning discharges.

Acknowledgement : With thank Dr.S.K. Mishra ,Principal, D.B.S. College, Kanpur for his helpful suggestions that improved this study .Thank are also acknowledged to Director N.P.L. New Delhi for providing the library facilities.

REFERENCES

1. M.A. Abbas, and J. Latham : The electrofreezing of supercooled water drops, J. Meteor. Soc. Japan, 47 (1969) 67-74.
2. F.F. Abraham : A re-examination of homogeneous nucleation theory: Thermodynamic aspects, J. Atmos. Sci., 25 (1968) 47-53.
3. J.T. Bartlett, A.P. van den Heuvel and B.J. Mason : The growth of ice crystals in an electric field, Z. Angew. Math. Phys., 14 (1963) 599-610.
4. F.C. Collins : Time lag in spontaneous nucleation due to non-steady state effects, Zeitschrift für Elektrochemie, 59 (1955) 404-407.
5. C.G. Deshpande and A.K. Kamra : Diurnal variations of the atmospheric electric field and conductivity at Maitri, Antarctica, J. Geophys. Res., 106 (D13) (2001) 14207-14218.
6. J.B. Doolittle and G. Vali : Heterogeneous freezing nucleation in electric fields, J. Atmos. Sci., 32 (1975) 375-379.
7. L. Dufour : Über das Gefrieren das Wassers und über die Bildung des Hagels, Poggendorfs, Ann. Physik, 114 (1861) 520-544.

8. L.F. Evans : The growth and fragmentation of the crystals in an electric field, *J. Atmos. Sci.*, 30 (1973) 1657-1664.
9. T.G. Gabarashvili and N.V. Gliki : Origination of the ice phase in supercooled water under the influence of electrically charged crystals, *Izv. Atmos. Oceanic Phys.*, 3 (1967) 570-574.
10. A.K. Kamra, C.G. Deshpande and Y. Gopalkrishnan : Challenge to the assumption of the unitary diurnal variation of the atmospheric electric field based on observations in the Indian Ocean, Bay of Bengal and Arbian Sea. *J. Geophys. Res.*, 99 (1994) 21043 - 21050.
11. A. Kantrowitz : Nucleation in very rapid vapour expansions, *J. Chem. Phys.*, 19 (1951) 1097-1100.
12. B. Karcher and U. Lohmann : A parameterization of cirrus cloud formation: Homogeneous freezing of supercooled aerosols, *J. Geophys. Res.*, 107 (D2) (2002) AAC.4-1 - AAC.4-10.
13. L.B. Loeb : A tentative explanation of the electric field effect on the freezing of supercooled water drops, *J. Geophys. Res.*, 68 (1963) 4475-4476.
14. T.C. Marshall and M. Stolzenburg : Electrical energy constraints on lightning, *J. Geophys. Res.*, 107(D7) (2002) ACL1-1 - ACL1-13.
15. B.J. Mason : *The Physics of Cloud*, Oxford University Press, London (1972).
16. J. Maybank and N.N. Barthakur : Growth and destruction of ice filaments in an electric field, *Nature*, 216 (1967) 50-52.
17. M. Mikki, V.A. Rakov, K.J. Rambo, G.H. Schnetzer and M.A. Uman : Electric fields near triggered lightning channels measured with Pockels sensors, *J. Geophys. Res.*, 107 (D16) (2002) ACL.2-1 - ACL.2-11.
18. G. Murino : Influence of electric fields on condensation of water vapour, *S. Afr. Tydskr. Fis.*, 2 (1979) 113-115.
19. R.F. Probstein : Time lag in the self nucleation of a supersaturated vapour, *J. Chem. Phys.*, 19 (1951) 619-626.
20. H.R. Pruppacher : The effect of an external electric field on the supercooling of water drops, *J. Geophys. Res.*, 68 (1963) 4463-4474.

21. H.R. Pruppacher : Electro-freezing of supercooled water, Pure Appl. Geophys., 104 (1973) 623-633.
22. W. Rau : Eiskeimbildung durch dielektrische polarization, Z.F. Naturforsch, 6a (1951) 649-657.
23. M. Roulleau, L.T. Evans, and N. Fukuta : The electrical nucleation of ice in supercooled clouds, J. Atmos. Sci., 28 (1971) 737-740.
24. W.D. Rust and R.J. Trapp : Initial balloon soundings of the electrical field in winter nimbostratus clouds in the U.S.A., Geophys. Res. Lett., 29 (2002) 20-1-20-4.
25. R.W. Salt : Effect of electrostatic field on freezing of supercooled water and insects, Science, 133 (1961) 458.
26. A.R. Sharma, N. Singh and S.D. Pandey : On the equivalence of external electric field to the supersaturation ratio in water vapour condensation, Ind. J. Radio & Space Phys., 21 (1992) 218-228.
27. N. Singh, J. Rai and N.C. Varshneya : The effect of external electric field on relaxation time in nucleation process of water vapour condensation and ice glaciation, Ann. Geophys., 42(1) (1986) 37-44.
28. M.H. Smith, R.F. Griffiths and J. Latham : The freezing of raindrops falling through strong electric fields, Quart. J. Roy. Meteorol. Soc., 97 (1971) 495-505.
29. H. Wakeshima : Time lag in the self nucleation, J. Chem. Phys., 22. (1954) 1614-1615.
30. P.K. Wang : The influence of atmospheric electricity on the precipitation, Seventh International Conference on Atmospheric Electricity, June 3-8, 1984. Albany, N.Y., U.S.A., 1984.

KĀTYĀYNA ŚULB-SŪTRA

VIRENDRA ARORA* AND NIDHI HANNA**

(Received 06.10.2004 and in revised form 06.10.2004)

ABSTRACT

Mature Science offers a new tool of interpretation to the historians who is then able to discover new results in those original works that they have already studied. Here Kandika 3 of Kātyāyana Śulba Sūtra deals with some geometrical figures and their transformations.

Key Words : KSS- Kātyāyana Śulba Sūtra, KSS-3/1 - first Verse of their Kandika of KSS.

INTRODUCTION

This paper aims at the conversion of area i.e. the conversion from rectangle into square and vice-versa ; circle in to square and vice-versa of various dimensions. It focuses how the four sided complicated structures have been simplified in the Śulba Sūtra. Here the endeavor has also been made to highlight the method of construction and proof of each sutra. This paper is discussed under the following topics.

- Extraction of a small square from large one.
- Conversion of a Rectangle into square.
- Transformation of a Lengthy rectngle in to a square.
- Transformation of a square in to rectangle.
- Some triplets for areas.
- Transformation of square in to circle and circle in to square.

EXTRACTION OF A SMALL SQUIRE FORM LARGE ONE

चतुरस्राच्चतुरसं निर्जिहीर्षन् यावन्निर्जिहीर्षे त्वावदुभयतोऽपच्छिद्य शाइकू

निखाय पाश्वमानी कृत्वा पाश्वमानी सम्मितामक्षण्यां तत्रोपसंहरति

समासेऽपच्छेदः सा करण्येष निहसि:

॥ 3/1 ॥

This means mark the length of the side of a small square on two opposite sides of a big one. Join the two points by a straight line Fix one point of straight line as center and other as moving point Extend the moving point to the base of big square. Now the large portion of base will be side of a required square which will be remaining area after subtracting the small square form the large one. Thus, small square, EFGH has been extracted form the given large square ABCD

* Dept. of Mathematics and Statistics, Gurukul Kangri University, Hardwar.

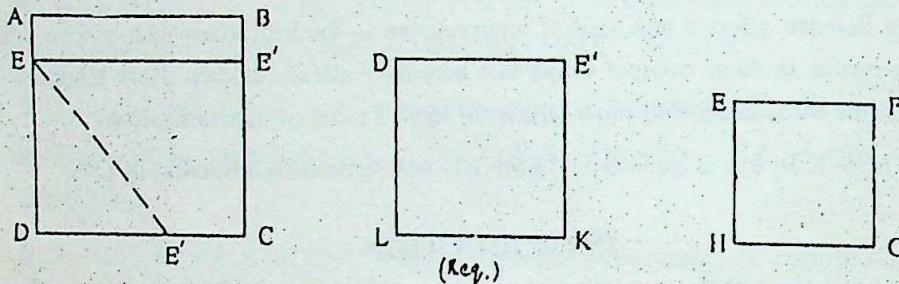
** Dept. of Mathematics and Statistics, Kanya Gurukul Mahavidyalaya, Hardwar

In right-angled triangle EDE'

$$ED^2 + DE'^2 = EE'^2$$

$$ED^2 + EF^2 = AB^2 \quad (EE' = AB, DE' = EF)$$

$$ED^2 = AB^2 - EF^2$$



CONVERSION OF A RECTANGLE INTO A SQUARE

दीर्घचतुरस्रं समचतुरस्रं चिकीर्षन्मध्ये तिर्यगपच्छिद्यान्यतरद्विभज्ये

तरत्पुरं-स्ताद्दक्षिणतश्चोपदध्याच्छेषमागन्तुना पूरयेत्स्योक्तो निर्हास

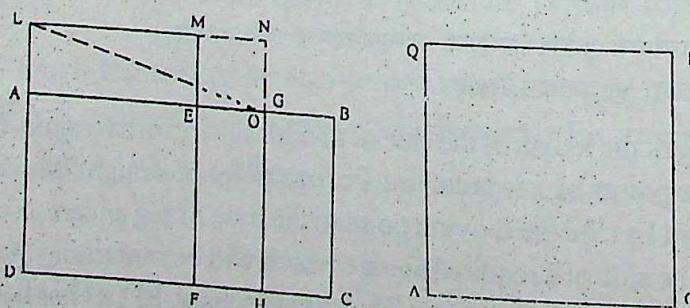
॥ 3/2 ॥

To Transfer a rectangle ABCD into square draw in line EF bisecting the lines AB and DC. Now in Sq. EBCF draw a line GH bisecting sides EB and FC. Place rect. GBCH at AE and we have LMEA. Now Complete a big square LNHD by adding sq. MNGE.

Mathematically

$$AO^2 = OL^2 - AL^2 \quad (OL = LN)$$

$$AO^2 = LN^2 - AL^2$$



Thus AO^2 represent the square AOPQ. Here square MNGE is being subtracted from the square LNHD to get the required square AOPQ.

TRANSFORMATION OF A LENGTHY RECTANGLE INTO A SQUARE

अतिदीर्घज्ये चिर्यड़ मान्याऽपच्छिद्यैक समासेन समस्य शेषं यथा योगमुपसहरे दिदत्येक समासः ॥ ३/३॥
A long rectangle ABCD is divided into equal squares ABOL, LOPM, MPQN and one small rectangle DCQN.

In fig. 2 a square will represent area of three squares (which is made on the hypotenuse of right angled triangle.)

Cut an arc $GT = FC$ in GR taking G as center. Join OT.

In fig. 3 a square made on OT will represent the req. area of given long rectangle.

$$\text{i.e. } AB^2 + LO^2 + MP^2 + NF^2 + FQ^2 = OT^2$$

$$\text{Since } AB^2 + LO^2 + MP^2 = OG^2$$

Again in $\triangle OGT$

$$OG^2 + GT^2 = OT^2$$

$$(GT = CF)$$

$$(CF^2 = FQ^2 + NF^2)$$

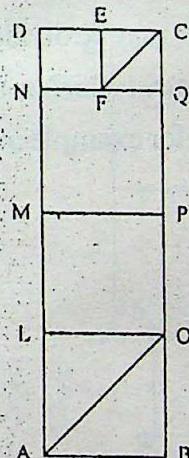


FIG. 1

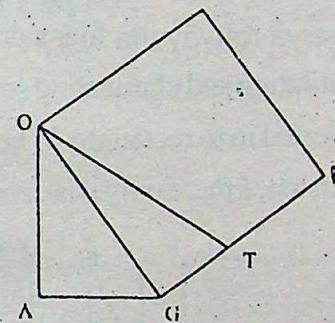


FIG. 2

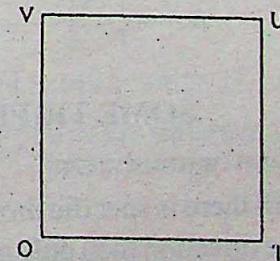


FIG. 3

Hence area of required square OTUV = area of rectangle ABCD

TRANSFORATION OF A SQUARE IN TO RECTANGLE

समचतुरसं दीर्घचतुरसं चिकीर्षमध्येऽक्षण्याऽपच्छिद्य विभज्येतरत्

पुरस्तादुत्तरतश्चोपदध्याद्, विषमं चेद् यथा-योगमुपसंहरेदिति व्यसः: ॥ 3/4 ॥

Divide a square by diagonal into two equal parts. Again divide one of the parts in to two so that we have two isosceles Δ s

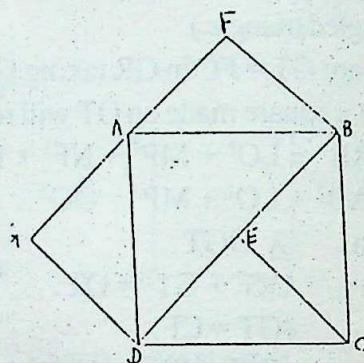
Place these two triangles on the opposite corners of the adjacent sides.

$$\Delta \text{ DEC} \equiv \Delta \text{ BEC}$$

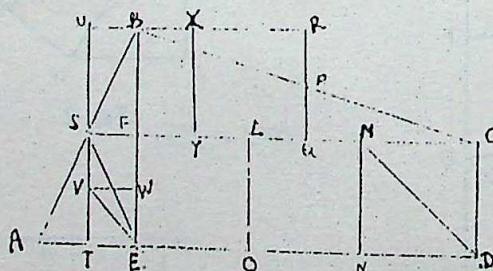
Place $\Delta \text{ BEC}$ on $\Delta \text{ FAB}$.

Similarly put $\Delta \text{ EDC}$ on $\Delta \text{ GDA}$.

GDBF is a required rectangle.



Some examples have been shown to describe this situation by Ācāryā Kātyāyana. According to Ācāryā Kātyāyana four sidea figure (quadrilateral) can be converted in to rectangle by substituting and eliminating some area. The figure may be divided in to small square then converted in to rectangle. Where the required figure (which is rectangle) do not come in to existence there only it is divided in to squares, for example in figure 1.



SOME TRIPLETS FOR AREA

प्रमाणं चतुरसमादेशादन्त्

॥ 3/5 ॥

Unless or untill there is specific direction of unit, a square unit is to be accepted. If there is some specific direction then only act accordingly.

द्वि प्रमाणा चतुःकरणी, त्रिप्रमाणा

नवकरणी, चतुः प्रमाणा षोडशकरणी

॥ 3/6 ॥

As much the length so much squared area will be got, because same digits will be multipiled with each other. If measure of unit is two then its area will be four unit of square,

If three then nine, if four then 16 and so on. Thus this is a little told about area.

यावत् प्रमाणा रज्जुर्भवति तावन्तस्तावन्तो

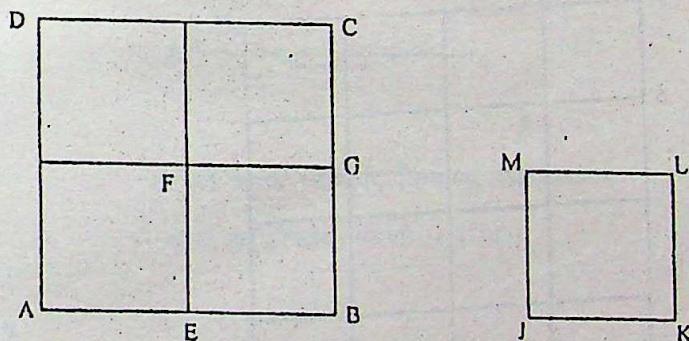
वर्गा भवन्ति तान्समस्येत्

॥ 3/7 ॥

As many units of measure in a chord so many squares of those units will be. Add all those squares then area will be got. If two units of measure taken, four squares will be formed for area ($2 \times 2 = 4$). If five units are there then 25 squares will be formed ($5 \times 5 = 25$). Similarly other can be produced.

अर्धप्रमाणेन पादप्रमाणं, विधीयते ॥ 3/8 ॥

But when half measure is taken then what will square of that space be ? If half the measure is squared it is fourth part of complete square $(1/2)^2 = 1/4$. If the measure is 24 units. Its half is 12. Square of 24 is 576 and that of 12 is 144. Thus 144 is one fourth of 576. This way for other units of measure can be applied.

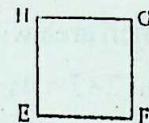
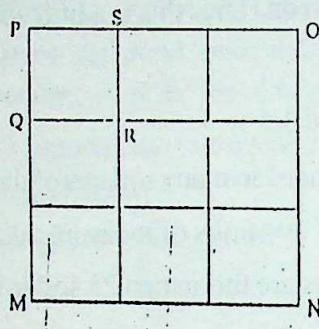


Example

Let us take ABCD whose measures is 24×24 sq. units. JKLM is a small square whose measure is 144 sq. units and this is fourth part of the square made by ABCD which is also equal to EBGF.

तृतीयेन नवमोऽशः ॥ 3/9 ॥

The square formed by one third the length of given measure is ninth part of total area of the given measure of unit. e.g. MP is given : 3 unit of measure, its third part is taken and square EFGH is completed. Now MNOP square is being completed and we see that PQRS is ninth part of MNOP. Area of this part PQRS is same as that of EFGH.

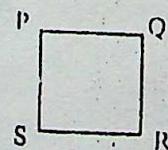
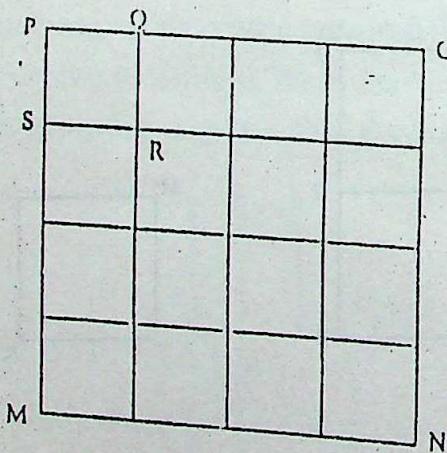


चतुर्थेन षोडशी कला

॥ 3/10 ॥

Area of the square which is one fourth of the length of a measure is sixteenth part of the square formed by full length of a measure. How can this area be separately deducted (nirhavasa).

[Consult Kss 3/1].



एष निर्हासस्तस्य पुरस्तादुक्तं शास्त्रम्

॥ 3/11 ॥

To deduct fourth part from 2 units of measure, ninth part from 3 units of measure, sixteenth part from 4 units of measure of side and so on, is already explained.

यावत्प्रमाणा रज्जुर्भवतीति विवृद्धे ह्रासाभवति

॥ 3/12 ॥

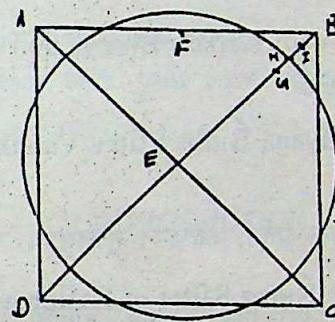
Deduction and entargement are made according to unit length of the chord. If lenght of chord is nine unit then its third part can be deducted. Similarly fourth part from sixteen units.

TRANSFORMATION OF SQUARE INTO CIRCLE & VICE-VERSA

चतुरस्रं मण्डलं चिकीर्षन्मध्यादसे निपात्य पाश्वतः परिलिख्य तत्र

यदतिरिक्तं भवति तस्य-तृतीयेन सह मण्डलं परिलिखेत्स समाधिः ॥ 3/13 ॥

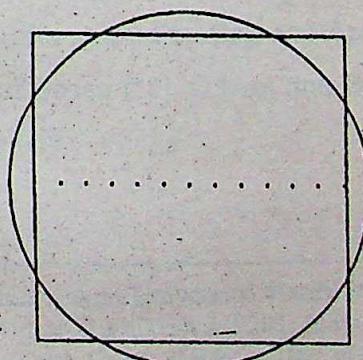
Transform ABCD square into a circle. Draw diagonal AC, BD from angle A and angle D. E is their meeting point, cut half of AB at F cut off on arc EG = FB from E. Divide GB into three parts GH, HI, IB. Now leave two parts and with centre EH draw a circle whose area will be equal to that square (for details see 2,p.65).



मण्डलं चतुरस्रं चिकीर्षन् विष्कम्भं पंचदशभागान्

कृत्वा द्वुद्धरेच्छेषः करणी ॥ 3/14 ॥

To convert a circle into square divide diameter into 15 parts.



Give up two parts out of them. With the help of remaining chord which is of 13 parts, draw a square. Now area of a square will be equal to that of the given circle (for details see 2, P 65).

After discussing the various śutras composed by Kātyāyana regarding the conversion of area it is concluded that in vedic age, there were various figures of different dimensions. It is also mentioned that in those days the device of dealing with the mathematical problems were very much in practice.

REFERENCES

1. Kātyāyana Śulba, edited with Karka's Bhasya and Mahidhara's vṛtti in the Kasi Sanskrit series, Banaras, 1936
2. Khadikar, S.D. (ed.), Kātyāyana Śulba Sūtra, Vaidika Saṃsodhana Mandala, Puna, 1974
3. R.C.D. Sharma, महाषिकात्यायनप्रणीतम्, कात्यायन शुल्बसूत्रम्, नाग प्रकाशक, दिल्ली (1993)
4. S. N. Sen & A.K. Bag, The Śulba Sūtra of Bañdhayana, Āpastamba, Kātyāyan and Manava, with text, English Translation and commentary, Indian National Science Academy, New Delhi, 1983.

New Coincidences and Fixed Points of Reciprocally Continuous and Compatible Hybrid Maps

S. L. SINGH* and AMAL M. HASHIM**

(Received 20.09.2004 and after revision 12.10.2004)

ABSTRACT

It is proved that a pair of reciprocally continuous and non-vacuously compatible single-valued and multivalued maps on metric spaces possesses a coincidence. This result is applied to obtain new general coincidence and fixed point theorems for single-valued and multivalued maps in metric spaces under tight minimal conditions.

Mathematics subject classifications (2000): 54H25, 47H10

Keywords and phrases: Coincidence point, fixed point, hybrid maps, reciprocally continuous, compatible maps.

INTRODUCTION

The notion of compatible maps, introduced by G. Jungck [7] (also called asymptotically commuting maps, cf. [24]) has proved useful for generalizing known results in the context of metric fixed point theory for continuous single-valued and multivalued maps (see, for instance, [1] - [2], [4], [7], [8], [11], [16], and [22]).

Recently, reciprocal continuity for a pair of (discontinuous) single-valued maps (resp. hybrid pair of maps) has been introduced in [16] (resp. [21]).

In this paper we obtain coincidence and fixed point theorems for a hybrid quadruple of maps on a metric space satisfying very general contractive type conditions which include several general conditions studied by Beg and Azam [1], Čirić [3], Jungck [8], Kaneko [10], [11], Kaneko and Sessa [12], Lohani and Badshah [13], Rhoades *et al.* [17], Tan and Minh [23] and others.

*Dept. of Maths & Stat. Gurukula Kangri University, Hardwar 249404

Mailing address: 21, Govind Nagar, Rishikesh 249201, India. E-mail: vedicmri@sancharnet.in.

**Permanent address: Dept. of Mathematics, College of Science, Univ. of Basrah, Iraq.

E-mail: amalmhashim@yahoo.com.

PRELIMINARIES

Consistent with [1], [14], [15] and [20], we will use the following notations, where (X, d) is a metric space and $CL(X)$ is a collection of all nonempty closed subsets of X . Further, $d(A, B)$ denotes the ordinary distance between nonempty subsets of A and B of X , while $d(x, B)$ stands for $d(A, B)$ when $A = \{x\}$. The distance function H is called the generalized Hausdorff metric for $CL(X)$ induced by the metric d of X .

Definition 2.1 (Singh and Mishra [21]). The maps $T: X \rightarrow CL(X)$ and $f: X \rightarrow X$ are reciprocally continuous (r. c.) on X (resp. at $t \in X$) if and only if $fTx \in CL(X)$ for each $x \in X$ (resp. $fTt \in CL(X)$) and $\lim_{n \rightarrow \infty} fTx_n = fM$, $\lim_{n \rightarrow \infty} Tfx_n = Tt$ whenever $\{x_n\}$ is a sequence in X such that

$$\lim_{n \rightarrow \infty} Tx_n = M \in CL(X), \lim_{n \rightarrow \infty} fx_n = t \in M.$$

For self-maps $f, g: X \rightarrow X$, this definition due to Pant [16] reads: f and g are r.c if and only if $\lim_{n \rightarrow \infty} gfx_n = gt$ and $\lim_{n \rightarrow \infty} fgx_n = ft$ whenever $\{x_n\} \subset X$ is such that $\lim_{n \rightarrow \infty} gx_n = \lim_{n \rightarrow \infty} fx_n = t \in X$. Clearly any continuous pair is reciprocally continuous but the converse is not true (Singh and Mishra [21, Example 2.3]). We remark that reciprocally continuous maps may be discontinuous even at their common fixed points (see [21]). For continuity of multivalued maps at their fixed and common fixed points, refer to [5].

The following definition is due to Kaneko and Sessa [12] and Beg and Azam [1] when $T: X \rightarrow CB(X)$.

Definition 2.2 (Singh and Mishra [21]). The maps $T: X \rightarrow CL(X)$ and $f: X \rightarrow X$ are compatible if and only if $fTx \in CL(X)$ for each $x \in X$ and $\lim_{n \rightarrow \infty} H(Tfx_n, fTx_n) = 0$, whenever $\{x_n\}$ is a sequence in X such that

$$\lim_{n \rightarrow \infty} Tx_n = M \in CL(X) \text{ and } \lim_{n \rightarrow \infty} fx_n = t \in M.$$

Evidently commuting maps T, f (i.e., when $fTx = Tfx, x \in X$) are weakly commuting (i.e., $H(Tfx, fTx) \leq d(fx, Tx), x \in X$, [10] and [19]), weakly commuting maps are compatible, and compatible T, f are weakly compatible (i.e., when $fTx = Tfx$, whenever $fx \in Tx$, see Jungck and Rhoades [9]), but the reverse implication is not true. For excellent discussion on the role of weak compatibility in fixed point considerations, refer to Jungck and Rhoades [op. cit.].

Definition 2.3 (Itoh and Takahashi [10]). The maps $T: X \rightarrow CL(X)$ and $f: X \rightarrow X$ are (IT)-commuting at a point $v \in X$ if $fTv \subset Tfv$ (cf. [21, page 629]).

Remark 2.4.

- (i) (IT)-commutativity of a hybrid pair (T, f) at a point v is more general than its compatibility (cf. [21, Example 1]) and weak compatibility at the point v .
- (ii) Continuous and commuting pair of hybrid maps (T, f) of a complete metric space need not have a coincidence; for example $Tx = \{1 + x\}$ and $fx = x, x \in [0, \infty)$, (see [21]).
- (iii) The nonvacuous compatibility of T and f implies the existence of at least a sequence $\{x_n\}$ in X such that

$$\lim_{n \rightarrow \infty} Tx_n = M \in CL(X) \text{ and } \lim_{n \rightarrow \infty} fx_n = t \in X. \text{ For details see [21].}$$

MAIN RESULTS

The following is the main result. In all that follows, $C(S, A)$ stands for the collection of coincidence points of S and A , that is, $C(S, A) = \{v; Sv \in Av\}$.

Theorem 3.1. Let (X, d) be a metric space and $A_i: X \rightarrow CL(X)$, $i = 1, 2, \dots$

and $S, T: X \rightarrow X$ such that

(i) $A_1X \subset TX$ and the pair (A, S) is reciprocally continuous and nonvacuously compatible,

(ii) $H(A_1x, A_1y) < (m_{1i}(x, y))$, $i > 1$ and $x, y \in X$ when $m_{1i}(x, y) > 0$,
where $(m_{1i}(x, y)) = \max \{d(Sx, A_1x), d(Ty, A_1y), d(Sx, A_1y),$
 $d(Ty, A_1x), d(Sx, Ty)\}$.

Then $C(S, A_1)$ and (T, A_i) are nonempty. Further,

(I) S and A_1 have a common fixed point Su , provided $SSu = Su$ for some $u \in C(S, A_1)$;

(II) T and A_i have a common fixed point Tw , provided $TTw = Tw$ and T, A_i are (IT)-commuting at $w \in C(T, A_i)$;

(III) S, T, A and A_i have a common fixed point, provided (i) and (ii) both are true.

Proof. Since A_1 and S are nonvacuously compatible, there exist a sequence $\{x_n\}$ in X such that $\lim_{n \rightarrow \infty} A_1x_n = M \in CL(X)$ and $\lim_{n \rightarrow \infty} Sx_n = u \in M$ and

$\lim_{n \rightarrow \infty} H(A_1Sx_n, SA_1x_n) = 0$. This, in view of r. c. of A_1 and S , yields $H(A_1u, SM) = 0$,

and $A_1u = SM$. Now $u \in M$ implies $Su \in A_1u$. Therefore $C(S, A_1)$ is nonempty.

Since $A_1X \subset TX$, there exists a point $w \in X$ such that $Su = Tw \in A_1u$.

If $A_1w \neq Tw$, then

$$\begin{aligned} d(Tw, A_iw) &\leq H(A_iu, A_iw) < \max \{d(Su, A_iu), d(Tw, A_iw), d(Tw, A_iu), \\ &\quad d(Su, A_iw), d(Tw, Su)\} \\ &= d(Tw, A_iw), \end{aligned}$$

a contradiction. This implies that $Tw \in A_iw$, and $C(T, A_i)$ is nonempty.

Further, $Su = SSu$ and the (IT)-commutativity of A and S at $u \in C(S, A)$ imply that $Su \in SA_iu = A_iSu$. So Su is a common fixed point of A and S .

The proof of (II) is similar, and (III) is immediate.

Corollary 3.2. Let (X, d) be a metric space, $A, B: X \rightarrow CL(X)$ and $S, T: X \rightarrow X$ such that

(iii) $AX \subset TX$ and the pair (S, A) is reciprocally continuous and nonvacuously compatible;

(iv) $H(Ax, By) \leq q(M(x, y))$ for each x, y in X and $q \in (0, 1)$, where

$$M(x, y) = \max \{d(Sx, Ax), d(Ty, By), d(Sx, By), d(Ty, Sx), d(Sx, Ty)\}.$$

Then $C(S, A)$ and $C(T, B)$ are nonempty. Further,

(Ia) S and A have a common fixed point Su , provided $SSu = Su$ for some $u \in (S, A)$;

(IIa) T and B have a common fixed point Tw , provided $TTw = Tw$ and T, B are (IT)-commuting at $w \in C(T, B)$;

(IIIa) S, T, A and B have a common fixed point, provided (Ia) and (IIa) both are true.

Proof. It may be completed following the proof of Theorem 3.1.

We remark that Theorem 3.1 with all maps single-valued presents a generalization of the main result of Pant [16, page 148]. Further, the above results extend some of the results of Singh and Kumar [18] to a quadruplet of hybrid maps.

Theorem 3.3. Corollary 3.2 with (iv) replaced by following :

(v) $H(Ax, By) < L(x, y)$ for all x, y in X when $L(x, y) > 0$, where

$$L(x, y) = \frac{\alpha d(Ty, By)[1 + d(Sx, Ax)]}{1 + d(Sx, Ty)} + \beta [d(Ty, By) + d(Sx, Ax)] \\ + \gamma [d(Sx, By) + d(Ty, Ax)] + \delta d(Sx, Ty),$$

where $\alpha, \beta, \gamma, \delta \geq 0$ and $\alpha + \beta + 2\gamma + \delta < 1$.

Proof. It may be completed following the proof of Theorem 3.1.

Remark 3.4. Theorem 3.3 generalizes a result of Lohani and Badshah [13] and others.

REFERENCES

1. Ismat Beg and Akbar Azam: Fixed points of asymptotically regular multivalued mappings, *J. Austral. Math. Soc. Ser. A* 53 (1992), 313-326. **MR1187851** (93m:54033).
2. T. H. Chang: Common fixed point Theorems for multivalued mapping, *Math. Japon.* 41(1995), No. 2, 311-320. **MR1326964** (96a:54057).
3. Lj. B. Čirić: A generalization of Banach's contraction principle, *Proc. Amer. Math. Soc.* 45 (1974), 267-273. **MR0356011** (50 #8484).
4. O. Hădăic: Common fixed point theorems for single and multivalued mappings, *Univ. Novom Sadu Zb. Rad. Prirod. Mat. Fak. Ser. Mat.*, 18(1988), No. 2, 145-151. **MR1100465** (92d:54060).

5. T. H. Hicks and B. E. Rhoades: Fixed points and continuity for multivalued mappings, *Internat. J. Math. Math. Sci.* 15(1992), 15-30. **MR1143924 (93b:54043)**.
6. S. Itoh and W. Takahashi: Single-valued mappings, multivalued mappings and fixed point theorems, *J. Math. Anal. Appl.* 59 (1977), 514-521. **MR0454752 (56#13000)**.
7. G. Jungck: Compatible mappings and common fixed points, *Internat. J. Math. Math. Sci.* 9 (1986), 771-779. **MR0939083 (89h:54029)**.
8. G. Jungck: Common fixed points for commuting and compatible maps on compacta, *Proc. Amer. Math. Soc.* 103 (1988), 977-983. **MR0947693 (89h:54030)**.
9. G. Jungck and B. E. Rhoades: Fixed points for set valued functions without continuity, *Indian J. Pure Appl. Math.* 29 (1998), 227-238. **MR1617919**.
10. H. Kaneko: A report on general contractive type conditions for multivalued mappings, *Math. Japon.* 33 (1988), 543-550. **MR0961199 (90a:47144)**.
11. H. Kaneko: A common fixed point of weakly commuting multi-valued mappings, *Math. Japon.* 33 (1988), no. 5, 741-744. **MR0972386 (90h:54055)**.
12. H. Kaneko and S. Sessa, Fixed point theorems for compatible multivalued and single-valued mappings, *Internat. J. Math. Math. Sci.* 12 (1989), 257-262. **MR0994907 (90i:54097)**.
13. P. C. Lohani and V. H. Badshah: Compatible mappings and common fixed point for four mappings, *J. Natur. Phy. Sci.* 11 (1997), 1-12. **MR1659342**.
14. S. B. Nadler, Jr.: Multivalued contraction mappings, *Pacific J. Math.* 30 (1969), 475-488. **MR0254828 (40 #8035)**.
15. S. A. Naimpally, S. L. Singh and J. H. M. Whitfield: Coincidence theorems for hybrid contractions, *Math. Nachr.* 127 (1986), 177-180. **MR0861724 (87m:54118)**.
16. R. P. Pant: A common fixed point theorem under a new condition, *Indian J. Pure Appl. Math.* 30 (1999), 147-152. **MR1681592**.
17. B. E. Rhoades S. L. Singh and Chitra Kulshrestha: Coincidence theorems for some multivalued mappings, *Internat. J. Math. Math. Sci.* 7 (1984), 429-434. **MR0771590 (86d:54073)**.
18. S. L. Singh, Ashish Kumar and Amal M. Hashim: Fixed point of contractive maps (2004), (reprint).
19. S. L. Singh, K. S. Ha and Y. J. Cho: Coincidence and fixed points of nonlinear hybrid contractions, *Internat. J. Math. Math. Sci.* 12 (1989), 247-256. **MR0994906 (90g:54042)**.

20. S. L. Singh and S. N. Mishra: Coincidences and fixed points of nonself hybrid contractions, *J. Math. Anal. Appl.* 256 (2001), 486-497. **MR1821752**.
21. S. L. Singh and S. N. Mishra: Coincidences and fixed points of reciprocally continuous and compatible hybrid maps, *Internat. J. Math. Math. Sci.* 30 (2002), 627-635. **MR1918077** (2003f:54102).
22. S. L. Singh and S. N. Mishra: On general hybrid contractions, *J. Austral. Math. Soc. Ser. A* 66 (1999), 244-254. **MR1671885** (2000b:54057).
23. D. H. Tan and N. A. Minh: Some fixed point theorems for mappings of contractive type, *Acta Math. Vietnam.* 3(1978), 24-42. **MR0527475** (80d:54054).
24. B. M. L. Tivari and S. L. Singh: A note on recent generalizations of Jungck contraction principle, *J. UPGC. Acad. Soc.* 3 (1986), 13-18. **MR0914072** (88k:54081).

COINCIDENCE AND FIXED POINT THEOREMS FOR HYBRID CONTRACTIONS

S. L. Singh* and Ritu Arora**

(Received 5.10.2004 and after revision 19.10.2004)

ABSTRACT

The purpose of this paper is to obtain coincidence and fixed point theorems for a pair of hybrid contracting maps.

Mathematics Subject Classifications (2000): 54H25, 54C60, 47H10.

Key words and phrases: Coincidence point, fixed point, hybrid contracting maps, hybrid contraction.

1. INTRODUCTION

Consistent with [1], [4], [11], [16], [18], [20] and [23], we will use the following notations, where (X, d) is a metric space. Let $CL(X)$ denote the collection of all nonempty closed subsets of X . The distance function H on $CL(X)$ is called the generalized Hausdorff metric induced by the metric d of X . Further, let $d(A, B)$ denote the ordinary distance between nonempty subsets A and B of X , while $d(A, x)$ stands for $d(A, \{x\})$ when B is the singleton $\{x\}$.

Let $T : X \rightarrow CL(X)$ be a multivalued map and $f : X \rightarrow X$ a single

* 21, Govind Nagar, Rishikesh 249201 India; Email: sh42@rediffmail.com

** Department of Holistic Medicine, Himalayan Institute Hospital Trust, Jolly Grant, Dehradun 248140

-valued map. Consider the following condition essentially due to Singh and Kulshrestha [21].

$$H(Tx, Ty) \leq q \max \{d(fx, fy), d(fx, Tx), (fy, Ty), \\ [d(fx, Ty) + d(fy, Tx)]/2\} \quad (\text{SK})$$

for all $x, y \in X$, where $q \in (0, 1)$.

We remark that (SK) with $f =$ the identity map on X is Ćirić's generalized multivalued contraction (Ćirić [4]), which in turn includes the well-known Nadler multivalued contraction (Nadler, Jr. [16], see also [1], [4], [9]-[13], [15], [23] and [25]).

The study of hybrid contracting maps involving single-valued and multivalued maps on metric spaces was initiated independently by Bhaskaran and Subrahmanyam [2], Hadžić [5], Kaneko [9], Kubiak [15], Naimpally [17] *et al.* and Singh and Kulshrestha [21]. (Here, according to Singh and Mishra [22], "hybrid contracting maps" means a pair of hybrid contraction or nonexpansive or contractive). Indeed, Singh and Kulshrestha [*op. cit.*] (see also [12], [19] and [22]) showed that T and f satisfying $T(X) \subseteq f(X)$ and (SK) have a coincidence, that is, there exists a point $z \in X$ such that $fz \in Tz$ when $f(X)$ is a complete subspace of X . This result is obviously true when, instead of $f(X)$, $T(X)$ is a complete subspace of X . For an immediate excellent generalization of this result one may refer to Rhoades *et al.* [18].

In all that follows, Y is an arbitrary nonempty set and (X, d) a metric space. Following Liu *et al.* [14], Singh and Mishra [23] and Tan *et al.* [26], we shall consider the following condition for $f: Y \rightarrow X$ and $T: Y \rightarrow CL(X)$.

$$H^2(Tx, Ty) \leq q \cdot m(x, y), \quad (1.1)$$

where $q \in (0, 1)$ and

$$\begin{aligned} m(x, y) := \max \{ & d^2(fx, fy), d(fx, fy) \cdot d(fx, Tx), d(fx, fy) \cdot d(fy, Ty), \\ & d(fx, fy) \cdot [d(fx, Ty) + d(fy, Tx)]/2, d(fx, Tx) \cdot d(fy, Ty), \\ & d(fx, Tx) \cdot [d(fx, Ty) + d(fy, Tx)]/2, d(fy, Ty) \cdot [d(fx, Ty) \\ & + d(fy, Tx)]/2, d(fx, Ty) \cdot d(fy, Tx) \}. \end{aligned}$$

Our main result is under the condition (1.1) (cf. Theorem 3). Following Rhoades *et al.* [18], we present generalized versions of this result as well.

2. RESULTS

DEFINITION 1. (Itoh and Takahashi [6], see also Singh and Mishra [24, p. 488]). Let Y be a nonempty set, $f: Y \rightarrow Y$ and $T: Y \rightarrow 2^Y$, the collection of all nonempty subsets of Y . Then the hybrid pair (T, f) is IT-commuting at $x \in Y$ if $Tfx \subseteq Tf^2x$ for each $x \in Y$. (This formulation in [25] is correct with the interchange of symbols for single-valued and multivalued maps).

We shall need the following result, which is a minor variant of a lemma due to Ćirić [4].

LEMMA 2. Let $A, B \in CL(X)$. Then for an $x \in A$ and for some q and k in $(0, 1)$, there exists a $y \in B$ such that

$$d^2(x, y) \leq q^k H^2(A, B).$$

THEOREM 3. Let Y be an arbitrary nonempty set and (X, d) a metric space. Let $T : Y \rightarrow CL(X)$ and $f : Y \rightarrow X$ be such that $T(Y) \subseteq f(Y)$ and (1.1) holds for all $x, y \in Y$. If $T(Y)$ or $f(Y)$ is a complete subspace of X then T and f have a coincidence. Indeed, for any $x_0 \in Y$, there exists a sequence $\{x_n\}$ in Y such that

- (I) $fx_{n+1} \in Tx_n$, $n = 0, 1, 2, \dots$;
- (II) $\{fx_n\}$ converges to fz for some $z \in Y$, and $fz \in Tz$, that is, T and f have a coincidence at z ; and
- (III) $d(fx_n, fz) \leq [\beta^n/(1-\beta)] \cdot d(fx_0, fx_1)$, where $\beta = q^{1-k}[1 + \sqrt{(1 + 8q^{1-k})}]$ for some $k \in (0, 1)$.

Further, if $Y = X$ and the pair (T, f) is IT-commuting at z then T and f have a common fixed point provided that $ffz = fz$.

PROOF. Pick a point x_0 in Y . Let k be a positive number such that $k < 1$. Following Singh and Kulshrestha [21] (see also [12]), we construct sequences $\{x_n\} \subseteq Y$ and $\{fx_n\} \subseteq X$ in the following manner.

Since $T(Y) \subseteq f(Y)$, we may choose a point $x_1 \in Y$ such that

$fx_1 \in Tx_0$. If $Tx_0 = Tx_1$ then $x_1 = z$ is a coincidence point of T and f , and we are done. So assume that $Tx_0 \neq Tx_1$ and choose $x_2 \in Y$ such that $fx_2 \in Tx_1$ and, by Lemma 2,

$$d^2(fx_1, fx_2) \leq q^{-k} H^2(Tx_0, Tx_1).$$

If $Tx_1 = Tx_2$, then x_2 becomes a coincidence point of T and f . If not, continue the process. In general, if $Tx_n \neq Tx_{n+1}$, we choose $fx_{n+2} \in Tx_{n+1}$ such that

$$d^2(fx_{n+1}, fx_{n+2}) \leq q^{-k} H^2(Tx_n, Tx_{n+1}).$$

Then by (1.1),

$$\begin{aligned} d^2(fx_{n+1}, fx_{n+2}) &\leq q^{-k} H^2(Tx_n, Tx_{n+1}) \\ &\leq q^{1-k} \max \{d^2(fx_n, fx_{n+1}), d(fx_n, fx_{n+1}).d(fx_n, Tx_n), d(fx_n, fx_{n+1}).d(fx_{n+1}, Tx_{n+1}), \\ &\quad d(fx_n, fx_{n+1}).[d(fx_n, Tx_{n+1}) + d(fx_{n+1}, Tx_n)]/2, d(fx_n, Tx_n).d(fx_{n+1}, Tx_{n+1}), \\ &\quad d(fx_n, Tx_n).[d(fx_n, Tx_{n+1}) + d(fx_{n+1}, Tx_n)]/2, d(fx_{n+1}, Tx_{n+1}).[d(fx_n, Tx_{n+1}) \\ &\quad + d(fx_{n+1}, Tx_n)]/2, d(fx_n, Tx_{n+1}).d(fx_{n+1}, Tx_n)\}. \end{aligned}$$

For the sake of simplicity, let $d_n := d(fx_n, fx_{n+1})$.

Then the above inequality, after simplification, reduces to

$$d_{n+1}^2 \leq q^{1-k} \cdot \max \{d_n^2, d_n d_{n+1}, d_n [d_n + d_{n+1}]/2, d_{n+1} [d_n + d_{n+1}]/2\}. \quad (3.1)$$

We remark that in the construction of sequences $\{x_n\}$ and $\{fx_n\}$, x_n (for each n) is not a coincidence point of T and f . This together with $Tx_n \neq Tx_{n+1}$ means that $fx_n \neq fx_{n+1}$. Indeed, if at any stage $fx_n = fx_{n+1}$ then

$fx_n \in Tx_n$, and x_n is a coincidence point of T and f . Therefore, according to our construction of the sequences, $d_n \neq 0$. Hence the inequality (3.1) implies one of the following:

$$d_{n+1}^2 \leq \lambda d_{n+1}^2, \quad \text{where } \lambda = q^{1-k};$$

$$d_{n+1} \leq \lambda d_n;$$

$$d_{n+1} \leq [\lambda/4 + \sqrt{(\lambda^2/16 + \lambda/2)}]d_n;$$

$$d_{n+1} \leq [\lambda/(2 - \lambda)]d_n.$$

Consequently,

$$d_{n+1} \leq \max \{ \sqrt{\lambda}, \lambda, [\lambda/4 + \sqrt{(\lambda^2/16 + \lambda/2)}], [\lambda/(2 - \lambda)] \} d_n.$$

This gives $d_{n+1} \leq \beta d_n$, when $\beta = [\lambda/4 + \sqrt{(\lambda^2/16 + \lambda/2)}]$. Notice that $0 < \beta < 1$. Hence $\{fx_n\}$ is a Cauchy sequence.

Now let $\mathcal{F}(Y)$ be a complete subspace of Y . Then the sequence $\{fx_n\}$ has a limit $f(Y)$. Call it b . Hence there exists a point $z \in Y$ such that $fz = b$. Since $d(fz, Tz) \leq d(fz, fx_{n+1}) + H(Tx_n, Tz)$, applying (1.1) to the last term of this inequality and making $n \rightarrow \infty$, the inequality gives $d(fz, Tz) \leq \sqrt{q} \cdot d(fz, Tz)$, and $fz \in Tz$, since $\sqrt{q} < 1$ and Tz is closed. This argument applies to the case when $T(Y)$ is a complete subspace of Y , since $T(Y) \subseteq \mathcal{F}(Y)$.

This proves (I) and (II). To see (III), let $m > n$. Then,

$$\begin{aligned} d(fx_n, fx_m) &\leq d(fx_n, fx_{n+1}) + \dots + d(fx_{m-1}, fx_m) \\ &\leq (\beta^n + \beta^{n+1} + \dots + \beta^{m-1}) \cdot d(fx_0, fx_1) \end{aligned}$$

$$< \beta^n / (1 - \beta) \cdot d(fx_0, fx_1).$$

Making $m \rightarrow \infty$, we get (III).

Let $\psi = \{\phi: \mathbb{R}_+ \rightarrow \mathbb{R}_+ | \phi \text{ is upper semicontinuous and nondecreasing}\}$. Then, in view of the proof of theorem 3 (above) and Rhoades *et al.* [18, Theorem 1], we have two generalizations of Theorem 3. We shall need the following definitions.

The following three definitions are due to Rhoades *et al.* [18] when $Y = X$.

DEFINITION 4. Let Y be an arbitrary nonempty set and (X, d) a metric space. If for a point $x_0 \in Y$ there exists a sequence $\{x_n\}$ in Y such that $fx_{n+1} \in Tx_n$, $n = 0, 1, 2, \dots$, then $O(T, f; x_0) := \{fx_n : n = 1, 2, \dots\}$ is the orbit for (T, f) at x_0 . Further $O(T, f; x_0)$ is called a regular orbit for (T, f) if for each n ,

$$d(fx_{n+1}, fx_{n+2}) \leq H(Tx_n, Tx_{n+1}).$$

We remark that the symbol for the orbit $O(T, f; x_0)$ is sometimes used for a sequence.

DEFINITION 5. A subspace $T(Y)$ or $f(Y)$ is called (T, f) -orbitally complete if every Cauchy sequence of the form $\{fx_{n_i} : fx_{n_i} \in Tx_{n_{i-1}}\}$ converges in X , where $\{x_n\}$ is a sequence in Y .

DEFINITION 6. If, for a point $x_0 \in Y$, there exists a sequence $\{x_n\}$ in Y such that the sequence $O(T, f; x_0)$ converges in X then X is called (T, f) -orbitally complete with respect to x_0 or simply $(T, f; x_0)$ -orbitally complete. Further, an orbit $O(T, f; x_0)$ is said to be asymptotically regular if $\lim_{n \rightarrow \infty} d(fx_n, fx_{n+1}) = 0$, where fx_n and fx_{n+1} are from $O(T, f; x_0)$.

THEOREM 7. Let Y be a nonempty set and (X, d) a metric space, $T : Y \rightarrow CL(X)$ and $f : Y \rightarrow X$ such that $T(Y) \subseteq f(Y)$ and the following hold:

$$H^2(Tx, Ty) \leq \phi(m(x, y)) \text{ for all } x, y \in Y; \quad (7.1)$$

$$\phi(t) < qt \quad \text{for each } t > 0, \quad (7.2)$$

for some $q \in (0, 1)$ and $\phi \in \psi$.

If $T(Y)$ or $f(Y)$ is a complete subspace of X then T and f have a coincidence point z (say). Further if $Y = X$ and the pair (T, f) are IT-commuting at z then T and f have a fixed point provided that $ffz = fz$.

THEOREM 8. Let Y be a nonempty set, (X, d) a metric space, $T : Y \rightarrow CL(X)$ and $f : Y \rightarrow Y$ such that

$$H^2(Tx, Ty) \leq \phi(M(x, y)), \quad (8.1)$$

where

$$\begin{aligned} M(x, y) := q \max \{ & d^2(fx, fy), d(fx, fy) \cdot d(fx, Tx), d(fx, fy) \cdot d(fy, Ty), \\ & d(fx, fy) \cdot d(fx, Ty), d(fx, fy) \cdot d(fy, Tx), d(fx, Tx) \cdot d(fy, Ty), \\ & d(fx, Tx) \cdot d(fx, Ty), d(fx, Tx) \cdot d(fy, Tx), d(fy, Ty) \cdot d(fx, Ty), \\ & d(fy, Ty) \cdot d(fy, Tx), d(fx, Ty) \cdot d(fy, Tx) \} \end{aligned}$$

for all $x, y \in Y$.

$$\phi(t) < t \text{ for each } t > 0, \quad \phi \in \psi. \quad (8.2)$$

If, for some $x_0 \in Y$, there exists a sequence $\{x_n\} \subseteq X$ such that the orbit $O(T, f; x_0)$ is regular and asymptotically regular, and $T(X)$ or $f(X)$ is $O(T, f; x_0)$ -orbitally complete, then T and f have a coincidence. Further, if $Y = X$ and the pair (T, f) is IT-commuting just at a coincidence point z (say) and $ffz = fz$, then fz is a common fixed point of T and f .

We remark that the completeness requirement in Theorem 3 may evidently be replaced by orbital completeness. Further, for the existence of a common fixed point of T and f , the requirement $ffz = fz$ can not be dropped from the above theorems (see, for instance, [17], [21], [22, page 194-195] and [24]). This means that, under either of the above theorems with $Y = X$, a coincidence point of T and f need not be their fixed point.

REFERENCES

1. Ismat Beg and Akbar Azam, Fixed points of asymptotically regular multivalued mappings, *Internat. J. Math. Math. Sci.* 15 (1992), 15-30.
2. R. Bhaskaran and P. V. Subrahmanyam, Common fixed points in metrically convex spaces, *J. Math. Phys. Sci.* 18 (5) (1984), 65-70.

3. H. W. Corley, Some hybrid fixed point theorems related to optimization, *J. Math. Anal. Appl.* 120 (1986), 528-532.
4. Lj. B. Ćirić, Fixed points for generalized multivalued contractions, *Math. Vesnik*, 9 (24) (1972), 265-272.
5. O. Hadžić, A coincidence theorem for multivalued mappings in metric spaces, *Studia Univ. Babes-Bolyai Math.* 26(1981), No. 4, 65-67.
6. S. Itoh and W. Takahashi, Single-valued mappings, multivalued mappings and fixed point theorems, *J. Math. Anal. Appl.* 59 (1977), 514-521.
7. J. Jachymski, On Reich's question concerning fixed points of multimap, *Boll. Un. Mat. Ital.* (7) 9-A (1995), 453-460.
8. G. Jungck and B. E. Rhoadeds, Fixed points for set-valued functions without continuity, *Indian J. pure Appl. Math.* 29 (1998), 227-238.
9. H. Kaneko, Single valued and multivalued f-contractions, *Boll. Un. Mat. Ital.* (6) 4-A (1985), 29-33.
10. H. Kaneko, A comparison of contractive conditions for multivalued mappings, *Kobe J. Math.* 3 (1986), 37- 45.
11. H. Kaneko and S. Seesa, Fixed point theorems for compatible multivalued and single valued mappings, *Internat. J. Math. Sci.* 12 (1989), 257-262.
12. C. Kulshrestha, Single-valued mappings, multi-valued mappings and fixed points theorems in metric spaces, Ph.D. Thesis (supervised by Prof. S. L. Singh), Garhwal University, Srinagar, 1983.

13. T. Kubiak, Two coincidence theorems for contractive type multivalued mappings, *Studia Univ. Babes-Bolyai Math.* 30 (1985), 65-68.
14. Z. Liu, F. Zhang and J. Mao, Common fixed points for compatible mappings of type (A), *Bull. Malaysian Math. Soc.* 22 (1999), 67-86.
15. S. N. Mishra, S. L. Singh and Rekha Talwar, Nonlinear hybrid contractions on Menger and Uniform spaces, *Indian J. Pure Appl. Math.* 25 (1994), 1039-1052.
16. S. B. Nadler, Jr., Multivalued contraction mappings, *Pacific J. Math.* 30 (1969), 475-488.
17. S. A. Naimpally, S. L. Singh and J. H. M. Whitfield : Coincidence theorems for hybrid contractions, *Math. Nachr.* 127(1986), 177-180.
18. B. E. Rhoades, S. L. Singh and C. Kulshrestha, Coincidence theorems for some multivalued mappings, *Internat. J. Math. Math. Sci.* 7(1984), 429-434.
19. K. P. R Sastry, I. H. N. Rao and K. P. R. Rao, A fixed point theorem for mutimaps, *Indian J. Phy. Natur. Sci.* 3 (sec B), (1983), 1-4.
20. S. L. Singh, K. S. Ha and Y. J. Cho, Coincidence and fixed points of nonlinear hybrid contractions, *Internat. J. Math. Math. Sci.* 12 (1989), 247-256.
21. S. L. Singh and Kulshrestha, Coincidence theorem in metric spaces, *Indian J. Phy. Natur. Sci.* 2B (1982), 19-22.
22. S. L. Singh and S. N. Mishra, Nonlinear hybrid contractions, *Journal of Natural & Physical Sciences* 5-8 (1993), 191-206.

23. S. L. Singh and S. N. Mishra, On general hybrid contractions, *J. Austral. Math. Soc. (Series A)* 66 (1999), 244-254.
24. S. L. Singh and S. N. Mishra, Coincidences and fixed points of nonself hybrid contractions, *J. Math. Anal. and Appl.* 256 (2001), 486-497.
25. S. L. Singh and S. N. Mishra, Coincidences and fixed points of reciprocally continuous and compatible hybrid maps, *Internat J. Math. Sci.* 30 (2002), 627-635.
26. Dejun Tan, Zeqing Liu and Jong Kyu Kim, Common fixed points for compatible mappings of Type (P) in 2-metric spaces, *Nonlinear Funct. Anal. & Appl.*, Vol. 8, NO. 2 (2003), 215-232.

प्राकृतिक एवं भौतिकीय विज्ञान शोध पत्रिका

खण्ड 18 अंक (2) 2004

(1) प्रकाशन स्थान : गुरुकुल कांगड़ी, विश्वविद्यालय, हरिद्वार

(2) प्रकाशन की अवधि : वर्ष में एक खण्ड अधिकतम दो अंक

(3) मुद्रक का नाम : चन्द्र किरण सैनी

राष्ट्रीयता : भारतीय

व पता : किरण ऑफसेट प्रिंटिंग प्रेस,

निकट गुरुकुल कांगड़ी फार्मसी,

कनरवल, हरिद्वार - 249404 फोन नं 0 245975

(4) प्रकाशक का नाम : डॉ ए०के० चोपड़ा

राष्ट्रीयता : भारतीय

व पता : कुलसचिव,

गुरुकुल कांगड़ी विश्वविद्यालय,

हरिद्वार - 249404

(5) प्रधान सम्पादक : डा० वीरेन्द्र अरोड़ा

राष्ट्रीयता : भारतीय

व पता : गणित विभाग,

गुरुकुल कांगड़ी विश्वविद्यालय,

हरिद्वार - 249404

(6) प्रबन्ध सम्पादक : डा० पी०पी० पाठक

राष्ट्रीयता : भारतीय

व पता : भौतिकी विभाग,

गुरुकुल कांगड़ी विश्वविद्यालय,

हरिद्वार - 249404

(7) स्वामित्व : गुरुकुल कांगड़ी विश्वविद्यालय,

हरिद्वार - 249404

मैं ए०के० चोपड़ा, कुलसचिव, गुरुकुल कांगड़ी विश्वविद्यालय, हरिद्वार, घोषित करता हूँ कि उपरिलिखित तथ्य मेरी जानकारी के अनुसार सही हैं।



हस्ताक्षर

डॉ ए०के० चोपड़ा
कुलसचिव

Aro



132081

GURUKUL KANGRI LIBRARY		
	Signature	Date
Access No.	Amrit	22/8/19
Class No.		
Cat No		
Tag etc		
E.A.R		
Recomm. By		
Data Ent. by	Amrit	22/8/19
Checked		

